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ANSWER KEY & SCHEME OF EVALUATION

B. Tech. (S3
Supplementary
April 2022

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REGULATION
2020

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Nadimpalli Satyanarayana Raju Institute of Technology (Autonomous), IQAC: Quality Management System (QMS)

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Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U.G.)	Program	Civil Engineering			Academic Year	2021 - 2022
Course Code	20CE304	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	Strength of Materials						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	State hook's Law.	20CE304.1	L1
2	What are the assumptions made in theory of simple bending?	20CE304.2	L1
3	What do you mean by moment area method?	20CE304.3	L1
4	Write the expression for crippling load when the both the ends of the column are hinged.	20CE304.4	L1
5	Differentiate between closed coil and open coil helical springs.	20CE304.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain stress – strain curve for mild steel rod.	6M	20CE304.1	L3
6 (b)	Derive the relations between Young's Modulus and Modulus of Rigidity.	6M	20CE304.1	L3
	OR			
7 (a)	A steel rod of 3 cm diameter and 5 m long is connected to two grips and the rod is maintained at a temperature of 95 degree Celsius. Determine the stress and pull exerted when the temperature falls to 30-degree C. (i) the ends do not yield (ii) the ends yield by 0.12 cm.	12M	20CE304.1	L3
8	Derive bending equation.	12M	20CE304.2	L3
	OR			
9 (a)	Derive shear stress distribution for circular section.	6M	20CE304.2	L3
9 (b)	Derive shear stress distribution for rectangular section.	6M	20CE304.2	L3
10 (a)	Explain conjugate beam method.	6M	20CE304.3	L3
10 (b)	Derive an expression for the slope and deflection of a Simply supported beam with a point load at center.	6M	20CE304.3	L3
	OR			
11	A beam of length 6m is simply supported at its ends and carries two point loads of 48 kN and 40 kN at a distance of 1m and 3m respectively from the left support. Find (i) Deflection under each load (ii). The point at which the maximum deflection occurs and Maximum deflection using Macaulay's method.	12M	20CE304.3	L3
12	Derive an expression for crippling load when both ends of the column are hinged.	12M	20CE304.4	L3
	OR			
13	A 1.75 m long steel column of rectangular cross-section 120 mm × 100 mm is rigidity fixed at one end and hinged at the other. Determine the buckling load on the column and the corresponding axial stress using Euler's formula. Take E for the column material as 200 GPa.	12M	20CE304.4	L3
14	Derive Torsion equation.	12M	20CE304.5	L4

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	OR			
15	The stiffness of a closely coiled helical spring is 1.5 N/mm of compression under a maximum load of 100N. The maximum shearing stress produced in the wire of the spring is 130 N/mm². The solid length of the spring (when the coils are touching) is given as 5cm. Find (i) Diameter of the wire (ii) Mean diameter of the coils and (iii) No. of coils required. Take C=4.5X10⁴ N/mm².	12M	20CE304.5	[illegible]

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Course : Strength of Materials
Program :- B.Tech. in Civil Engineering
Course code : 20CE304.

Max Marks = 70
Academic Year = 2021-2022.
Semester III

PART- A

1. State hook's law.

Ans: Hook's law states that when the material is loaded within elastic limit, the stress is proportional to the strain produced by the stress.

2. What are the assumption made in theory of simple bending

Ans: Assumptions:-

- The material of the beam is homogenous and isotropic
- The value of young's modulus of elastic is the same in tension and compression
- The transverse sections which were plane before bending remain plane after bending.
- The beam is initially straight and all longitudinal filaments bend into circular arc with common centre of curvature.

①

→ The radiu of curvature is Large compared with the dimension of the cross section

→ Each Layer of the beam is free to expand or contract, independently of the layer, above or below it.

3. What Do you Mean by Moment Area method?

An: <u>Moment Area Method:</u>

The Moment Area method uses the area of moment divided by the flexural rigidity ($\frac{M}{EI}$) diagram of the beam to determine the slope and deflection along the beam.

4. Write the expression for crippling load when the both ends of the column are hinged?

An: The load at which the column just buckles (or bends) is called crippling load.

Consider a column AB of length "l" and uniform cross sectional area, hinged at both of its ends, A and B. Let P be the crippling load at which the column has just buckled.

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crippling load when both ends of the column are hinged

$$\boxed{P = \frac{\pi^2 EI}{l^2}}$$ l actual length.

$$P = \frac{\pi^2 EI}{l_e}$$ for effective length.

r/b effective length and actual length $l_e = l$.

5) Differentiate between closed coil and open coil helical springs?

Ans:

Helical springs

- Closed coiled helical springs
- open coiled helical springs.

closed coiled helical springs	open coiled helical springs
- closed coiled helical springs are the springs in which helix angle is very small	- wire of the open coiled helical spring is wound not so tightly and thus sufficient space a gap each b/w two adjacent coils
- or The pitch b/w two adjacent turns is small	
- The closed helical springs carrying axial load	- pitch of spring wire is comparatively large e a resultant large helix angle
- The bending effect Removed.	
- It consists of purely torsional stress	- It subjected to axial load and exten^ compre.

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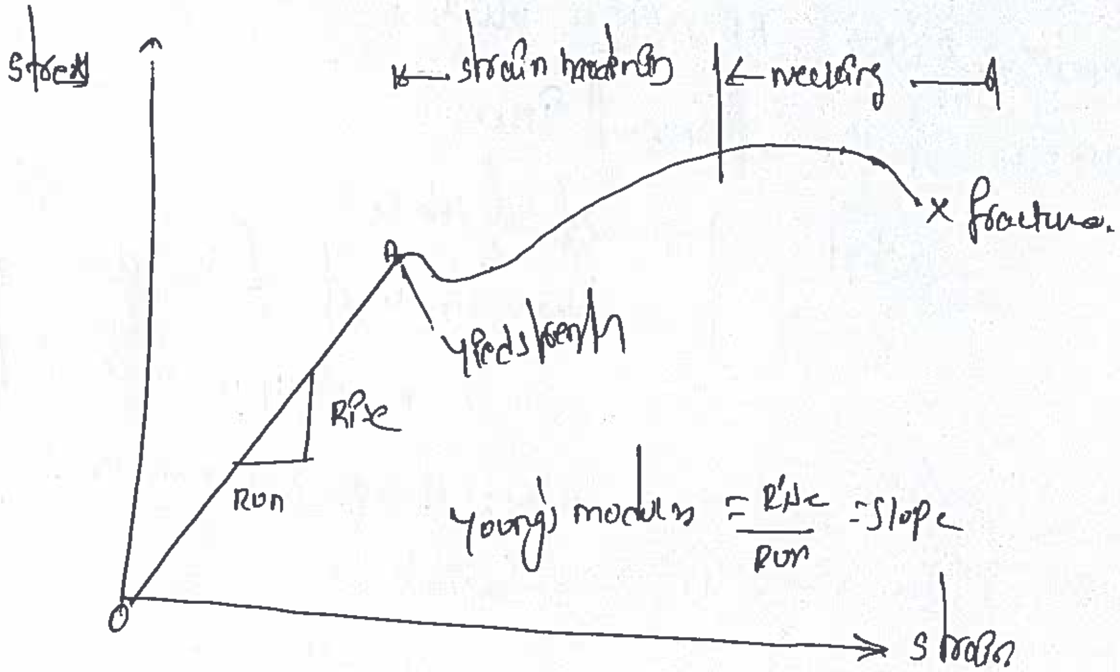
PART-B

6a) Explain stress – strain curve for mild steel rod.

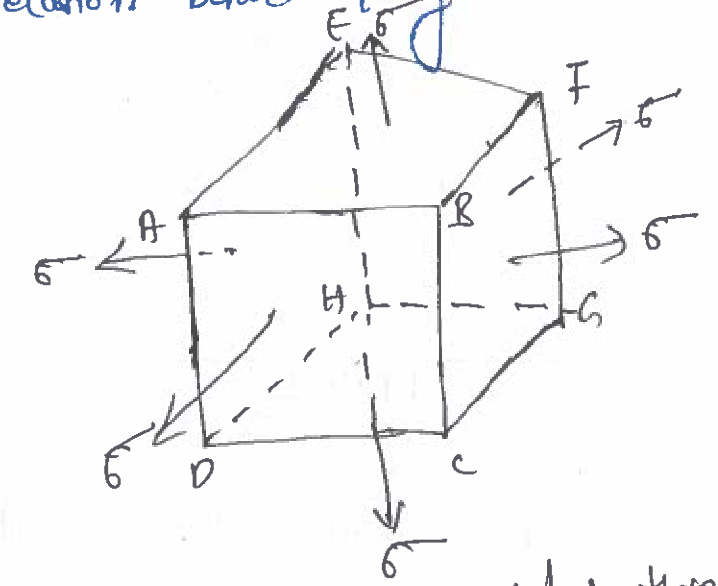
Ans:



OA = limit of proportionality

6b) Derive the relation between young's Modulus and Modulus of rigidity?

Ans'



from fig, a cube ABCDEFGH which is subjected to three mutually perpendicu Tensile stresses of equal intensity,

Let L = length of cube
dL = change in length of the cube
E = young's modulus of the material of the cube
σ = Tensile stress acting on the faces.
μ = Poisson ratio.

The volume of the cube = V = L³.

→ The strain (AB) of the side of the cube. three mutually perpendicular stresses.

→ The following three strains:—

① strain AB due to stresses on the faces AEHD.

② strain AB due to stresses on the faces AEFB and DHGC.

The compressive lateral strain is equal to $-\mu \frac{\sigma}{E}$.

→ strain AB due to stresses on the faces ABCD and EFGH, This also compressive lateral strain and is equal to $-\mu \frac{\sigma}{E}$.

Total strain AB is given by

$$\frac{dL}{L} = \frac{\sigma}{E} - \mu \times \frac{\sigma}{E} - \mu \times \frac{\sigma}{E}$$

$$= \frac{\sigma}{E}(1-2\mu) \rightarrow (i)$$

original volume $(V) = L^3$.

If dL is the change in length, then dV is the change in volume.

Differentiation equation (ii) with respect L.

$$dV = 3L^2 \times dL \rightarrow (iii)$$

Dividing equation (iii) by equation (ii)

$$\frac{dV}{V} = \frac{3 \times L^2 \times dL}{L^3} = \frac{3dL}{L}.$$

Substitute the value of $\frac{dL}{L}$ from equation (i), the above equation we get

$$\frac{dv}{v} = \frac{\sigma}{\left(\frac{dv}{v}\right)} = \frac{\sigma}{\frac{3\sigma}{E}(1-2\mu)}$$

$$= \frac{E}{3(1-2\mu)}$$

(4)

$$\boxed{E = 3K(1-2\mu)}$$

7(a) A steel rod of 3cm diameter and 5m long is connected to two grips and the rod is maintained at a temperature of 95° celsius. Determine the stress and pull exerted when the temperature falls to 30°C.

① the ends donot yield

(ii) the ends yield by 0.12cm. Take $E = 2\times10^5$ MN/m^2 and $\alpha = 12\times10^{-6}/°C$.

Ans:- Given :-

Dia of the rod $d = 3cm = 30mm$

Area of the rod, $A = \frac{\pi}{4} \times 30^2 = 225\pi$ mm^2

Length of the rod $L = 5m = 5000mm$

Initial temperature $T_1 = 95°C$

Final temperature $T_2 = 30°$

Fall in temperature $T = T_1 - T_2 = 95-30 = 65°C$.

Modulus of elasticity $E = 2\times10^5$ MN/m^2

$= 2\times10^5 \times 10^6$ N/m^2

$= 2\times10^6$ N/m^2

Coefficient of linear expansion $\alpha = 12\times10^{-6}/°C$.

a) when the ends do not yield:-

stress $= \alpha . T . E$

$= 12\times10^{-6} \times 65 \times 2\times10^{11}$ N/m^2.

$= 156$ N/mm^2 (tensile

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Pull in the rod = Stress x Area

$= 150 \times 22.5\pi$

$= 110269.9N$

(ii) when the ends yield by 0.12 cm

$\delta = 0.12cm = 1.2mm$,

The stress when the ends yield is given by equation.

$$\text{stress} = \frac{\alpha T L - \delta}{L} \times E$$

$$= \frac{12 \times 10^{-6} \times 65 \times 5000 - 1.2}{5000} \times 2 \times 10^5 \text{ N/mm}^2$$

$$= \frac{(3.9 - 1.2)}{5000} \times 2 \times 10^5 \text{ N/mm}^2.$$

$= 108 \text{ N/mm}^2$.

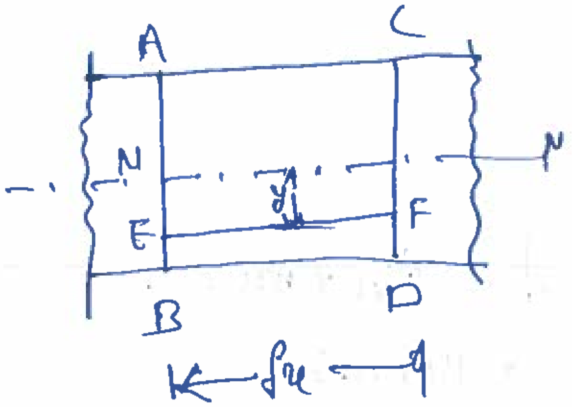
Pull in the rod = Stress x Area

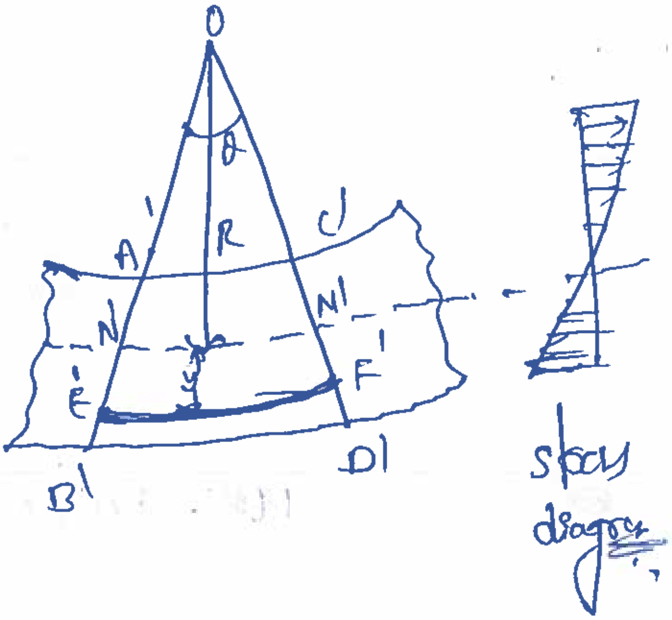
$= 108 \times 225\pi = 76340.7N$

⑤

6) Derive bending equation?

A).





consider, a small length δx of a beam subjected to simple bending due to the action of bending, the part of length δx will be deformed as shown in fig.

Let A'B' and C'D' meet at "O".

Let R = Radius of neutral layer N'N'

θ = angle subtended at O by A'B' and C'D' produced

strain variation along the ~~length of~~ depths of the beam

consider a layer EF at a distance y below the neutral layer NN. After bending the layer will be elongated to E'F';

original length of layer EF = δx,

length of the neutral layer NN = δx.

After bending, the length of Neutral layer N'N' will remain unchanged, but length of layer E'F' will increase.

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now from fig, - $N'N' = R \times \theta$

and $E'F' = (R+y)\theta$ (radius of $E'F' = R+y$.

But $N'N' = NN = \delta x$

Hence $\delta x = R \times \theta$.

Increase in the length of the layer EF

$= E'F' - EF = (R+y)\theta - R\times\theta$.

$= y \times \theta$.

Strain in the layer EF

$$= \frac{\text{Increase in the length}}{\text{original length}}$$

$$= \frac{y\times\theta}{EF} = \frac{y\times\theta}{R\times\theta} \quad (\because EF = \delta x = R\times\theta)$$

$= y/R$.

As R is constant. hence the strain in layer is proportional to its distance from the neutral axis. the above equation shows the variation of strain along the depth of the beam the variation of strains linear. ⑥

Stress variation:-

Let σ = stress in the layer EF

E = young's modulus of the beam.

then $E = \dfrac{\text{stress in the layer EF}}{\text{strain in the layer EF}}$

$= \dfrac{\sigma}{(y/R)}$ (∵ strain in EF = y/R)

$$\boxed{\sigma = E \times \frac{y}{R} = \frac{E}{R} \times y}$$

Since E and R are constant. therefore stress in any layer is directly proportional to the distance of the layer from the neutral layer.

∴ The variation of stress along the depth of the beam, the variation of stress is linear.

$$\boxed{\frac{\sigma}{y} = \frac{E}{R}}$$

Neutral axis and Moment of Resistance:- The neutral axis of any transverse section of a beam is defined as the line of intersection of the neutral layer with the transverse section. It is written as N.A.

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- If a section of a beam is subjected to pure sagging moment, then the stresses will be compressive at any point above the neutral axis and tensile below the neutral axis. There is no stress at the neutral axis.
- The stress at a distance y from the neutral axis is given by

$$\boxed{\sigma = \frac{E}{R} \times y}$$

- The cross section of a beam, let N.A be the neutral axis of the section. Consider a small layer at a distance y from the neutral axis.

Let dA = Area of the layer.

Now force on the layer
= stress on layer × Area of layer

$$= \sigma \times dA$$

$$= \frac{E}{R} \times y \times dA \rightarrow ① \quad \left(\sigma = \frac{E}{R} \times y\right)$$

- Total ~~section on~~ force for the beam section is obtained by
Integrating the above equation
total force on the beam section

Ⓟ

$$\frac{E}{R}\int y\times dA=0$$

or

$$\int y\times dA=0$$

Moment of resistance: Due to pure bending, the layers above the N.A are subjected to compressive stress whereas the layers below the N.A are subjected to tensile stress.

∴ These force will have moment about the N.A.

Force on layer $=\frac{E}{R}\times y\times dA$.

Moment of this force about N.A

$=$ Force on layer $\times y$

$=\frac{E}{R}\times y\times dA\times y$

$=\frac{E}{R}\times y^2\times dA$.

∴ Total moment of the forces on the section of the beam

$=\int\frac{E}{R}\times y^2\times dA=\frac{E}{R}\int y^2\times dA$.

Let M = External moment applied on the beam section. For equilibrium the moment of resistance offered by the section should be equal to the external bending moment

$$\boxed{M=\frac{E}{R}\int y^2\times dA}$$

→ Let M = External moment applied on the beam section. For equilibrium the moment of resistance offered by the section should be external bending moment.

$$\boxed{M = \frac{E}{R}\int y^2 \times dA}$$

→ But the expression $\int y^2 \times dA$ represents the moment of Inertia of the area of the section about the neutral axis.

Let the Moment of Inertia be I.

$$M = \frac{E}{R} \times I \text{ or } \frac{M}{I} = \frac{E}{R}$$

we have $\frac{\sigma}{y} = \frac{E}{R}$

$$\boxed{\frac{M}{I} = \frac{\sigma}{y} = \frac{E}{R}}$$

M is expressed in Nmm ; I in mm⁴

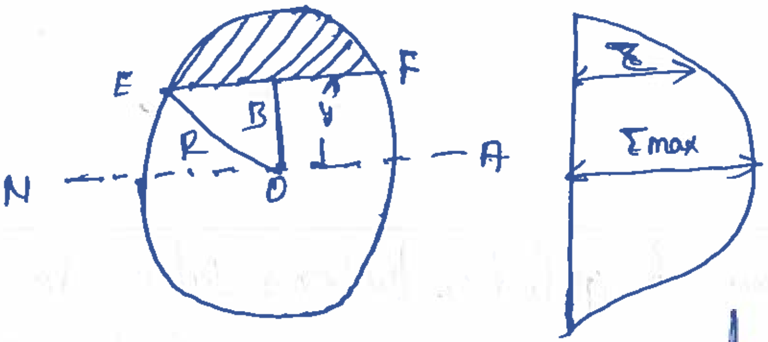
σ is expressed in N/mm² ; y in mm

E is expressed in N/mm² , R in mm

⑧

9a) Derive shear stress distribution for circular section.

Ans.



From Circular Section: Let R is the radius of circular section, of F

is the shear force acting on the section

→ consider a level EF at a distance y from the neutral axis

→ The shear stress at this level is given by.

$$\boxed{\tau = \frac{F \times A\bar{y}}{I \times b}}$$

where $A\bar{y}$ = moment of the shaded Area about the neutral axis (N.A)

I = Moment of Inertia of the whole circular section

b = width the beam at the level EF.

consider a strip thickness dy at a distance y from N.A

Let dA is the area of strip.

$$dA = b \times dy = EF \times dy$$

$$= 2 \times EB \times dy$$

$$= 2 \times \sqrt{R^2 - y^2} \times dy$$

$= 2 \times EB \times dy$

$= 2 \times \sqrt{R^2-y^2} \times dy$ [

side $EB = \sqrt{R^2-y^2}$)

→ Moment of this area about N.A

$= y \times dA$

$= y \times 2\sqrt{R^2-y^2} \times dy$

$= 2y\sqrt{R^2-y^2}\, dy$.

→ Moment of the whole shaded area about N.A is obtained by Integration the above equation b/w the limits. y and R.

$$A\bar{y} = \int_y^R 2y\sqrt{R^2-y^2}\, dy$$

$$= -\int_y^R (-2y)\sqrt{R^2-y^2}\, dy.$$

now (−2y) is the differential of (R^2-y^2).

Integration of the above equation becomes

$$A\bar{y} = -\left[\frac{(R^2-y^2)^{3/2}}{3/2}\right]_y^R$$

Ⓣ

$$= -\frac{2}{3}\left[(R^2 - R^2)^{3/2} - (R^2 - y^2)^{3/2}\right]$$

$$= -\frac{2}{3}\left[0 - (R^2 - y^2)^{3/2}\right] = \frac{2}{3}\left[R^2 - y^2\right]^{3/2},$$

Substituting the value of $A\bar{y}$ in equation ① we get

$$q = \frac{F \times \frac{2}{3}(R^2 - y^2)^{3/2}}{I \times b}$$

But $b = EF = 2 \times EB = 2 \times \sqrt{R^2 - y^2}$.

Substituting the value of b in the above equation

$$q = \frac{\frac{2}{3}F(R^2 - y^2)^{3/2}}{I \times 2 \times \sqrt{R^2 - y^2}} = \frac{F}{3I}(R^2 - y^2) \rightarrow ③$$

From equation ③ At $y = 0$ at neutral axis, the shear stress is maximum

$$q_{max} = \frac{F}{3I} R^2$$

$$I = \frac{\pi}{64} D^4 = \frac{\pi}{64} \times (2R)^4 \qquad D = 2R$$

$$= \frac{\pi}{4} R^4$$

$$q_{max} = \frac{F \times R^2}{3 \times \frac{\pi}{4} \times R^4} = \frac{4}{3} \times \frac{F}{\pi R^2}$$

average shear stress $q_{av} = \frac{\text{shear force}}{\text{Area of cross section}} = \frac{F}{\pi R^2}$

$q_{max} = 4/3 \times q_{av}$

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9b) Derive shear stress distribution for rectangular section?

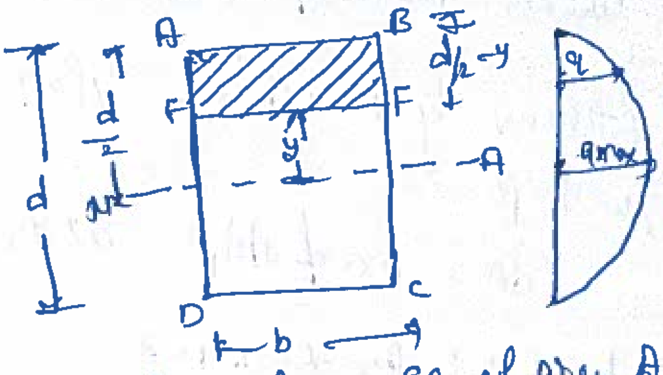
Ans: Rectangular section of beam of width b and depth d. let F is the shearforce acting at the section.

Consider a level EF at a distance y from the neutral axis.

The shear stress at this level given by equation.

$$\boxed{\tau = F \times \frac{A\bar{y}}{I b}}$$

where A = Area of the section above (Shaded Area ADEF)



$\bar{y}$ = Distance of the C.G of area A from neutral axis

$$= y + \frac{1}{2}\left(\frac{d}{2} - y\right) = y + \frac{d}{4} - \frac{y}{2} = \frac{y}{2} + \frac{d}{4} = \frac{1}{2}\left(y + \frac{d}{2}\right)$$

b = Actual width of the section at the level EF

I = MOI of the whole section about N.A.

Substitute these values in the above equation.

$$\tau = \frac{F \times \left[\frac{d}{2} - y\right] \times b \times \frac{1}{2}\left(y + \frac{d}{2}\right)}{b \times I}$$

⑩

$$= \frac{F}{2I}\left[\frac{d^2}{4} - y^2\right]$$

At top edge, $y = d/2$ and hence

$$\tau = \frac{F}{2I}\left[\frac{d^2}{4} - \left(\frac{d}{2}\right)^2\right] = \frac{F}{2I} \times 0 = 0$$

At the neutral axis, $y = 0$ hence.

$$\tau = \frac{F}{2I}\left[\frac{d^2}{4} - 0\right] = \frac{F}{2I} \times \frac{d^2}{4}$$

$$= \frac{Fd^2}{8I} = \frac{Fd}{8 \times \frac{bd^3}{12}}$$

$$= \frac{12}{8} \times \frac{F}{bd} = 1.5 \times \frac{F}{bd}.$$

Now Average shear stress $\tau_{av} = \frac{\text{Shear force}}{\text{Area of section}} = \frac{F}{b \times d}$

$$\boxed{\tau = 1.5 \times \tau_{avg}}$$

$\tau = \frac{A\bar{y}}{Ib}$ where $A\bar{y}$ = moment of shaded area about N.A

Consider a strip of thickness dy at a distance y from N.A. Let dA is the area of this strip.

$$dA = \text{Area of strip} = b \times dy$$

Moment of shaded area about N.A

$$= \int_y^{d/2} y \times b \times dy$$

$$= b\int_y^{d/2} y \times dy \quad \text{(as b is constant)}$$

$$= b\left[\frac{y^2}{2}\right]_y^{d/2} = \frac{b}{2}\left[\left(\frac{d}{2}\right)^2 - y^2\right]$$

$$A\bar{y} = b/2\left[\frac{d^2}{4} - y^2\right] = b/2\left[d^2/4 - y^2\right]$$

$$\tau = \frac{F \times b/2 \times (d^2/4 - y^2)}{I \times b} = \frac{F}{2I}\left(d^2/4 - y^2\right)$$

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10a) Explain conjugate beam method?

Ans: The slopes and deflection of beams can obtained by conjugate beam method.

Conjugate beam method: Before describing the conjugate beam method, we first define conjugate beam.

- Conjugate is a imaginary beam of length equal to the that of the original beam but for which the load diagram is the $\frac{M}{EI}$ diagram. (the load at any point on the conjugate beam is also equal to BM at the point divided by EI.
- The slopes and deflection at any section of beam by conjugate beam method is given,

① The slope at any section of the given beam is equal to the shear force at the corresponding section of the conjugate beam.

② The deflection at any section for the given beam is equal to bending moment at the corresponding of the conjugate beam.

- Before applying the conjugate beam method, conjugate beam is constructed. The load on the conjugate beam at any point is equal to the B.M. at that point divided by EI. Hence the loading on the conjugate beam is known.

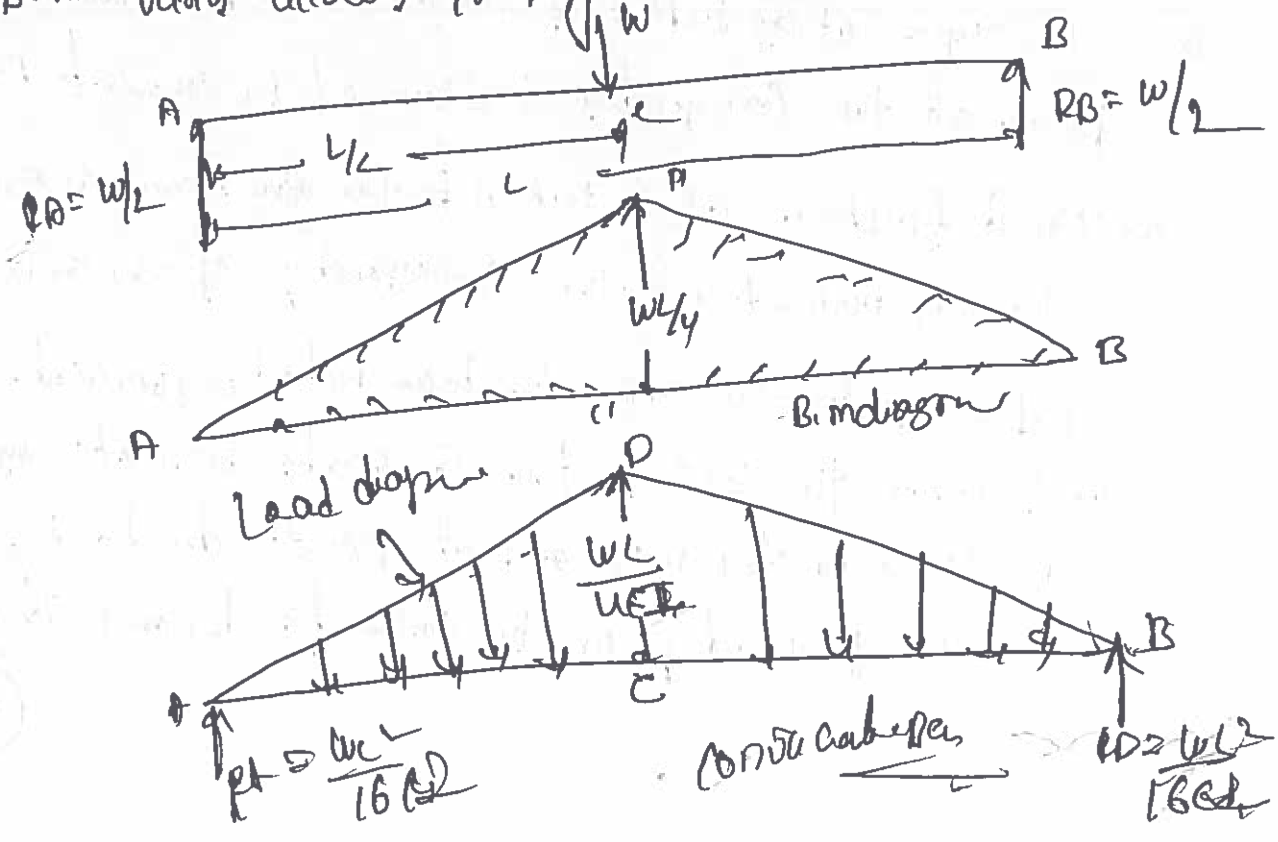
⑪

- The Shearforce at any Point on the conjugate beam gives the slope at the corresponding point of actual beam.
- The B.m at any Point gate beam give the slope at the corresponding point of actual beam.
- The B.m at any Point on the conjugate beam gives the deflection at the corresponding point of the actual beam.
- M/EI diagram is diagram which shows the variation of M/EI over the length of the beam.

(0b) Derive an expression for the slope and deflection of simply supported beam with point load at centre?

Ans: A simply supported Beam AB of length L carrying point load at W at the centre C. The B.M at A and B is zero and at the centre B.M will be $WL/4$.

- The B.M varies according to straight line law.



→ The B.M varies according to straight line law. The B.M diagram is shown in fig.

→ Now the conjugate beam AB can be constructed. The load on the conjugate beam will be obtained by dividing the B.M at that point by EI.

→ The shape of loading on the conjugate beam will be same as of B.M diagram.

→ The ordinate of loading on conjugate beam will be equal to

$$\frac{M}{EI} = \left(\frac{\frac{WL}{4}}{EI}\right) = \frac{W \times L}{4EI}$$

. Hence ordinate at the centre will be $\frac{WL}{4EI}$ as shown in fig.

→ Finally, let R_A = Reaction at A for conjugate beam

R_B = Reaction at B for conjugate beam.

Total load on the conjugate beam

= Area of the load diagram

$$= \frac{1}{2} \times AB \times CD = \frac{1}{2} \times l \times \frac{WL}{4EI}$$

$$= \frac{WL^2}{8EI}$$

→ Reaction at each support for the conjugate beam will be half of the total load.

$$RA = RB = \frac{1}{2} \times \frac{WL^2}{8EI} = \frac{WL^2}{16EI}$$

Let θ_A = slope at A for the given beam $\left(dy/dx\right)$ at A.

y_c = deflection at C for the given beam.

then According to conjugate beam method.

θ_A = Shear force at A for the conjugate beam

= RA (S.F at for conjugate beam = RA)

$$= \frac{WL^2}{16EI}$$

as y_c = B.M at C for the conjugate beam -

$= RA \times \frac{L}{2}$ - load corresponding to ACD

× Distance of CG of ACD from C

$$= \frac{WL^2}{16EI} \times \frac{L}{2} - \left(\frac{1}{2} \times \frac{L}{2} \times \frac{WL}{4EI}\right) \times \left(\frac{1}{3} \times \frac{L}{2}\right)$$

$$= \frac{WL^3}{32EI} - \frac{WL^3}{96EI} = \frac{3WL^3 - WL^3}{96EI}$$

$$\boxed{y = \frac{WL^3}{48EI}}$$

ANSWER KEY AND SCHEME OF EVALUATION

4) A beam of length 6m is simply supported at its ends and carries two point loads of 48kN and 40kN at a distance of 1m and 3m respectively from the left support? $E = 2\times105\ N/mm^2$, $I = 85\times10^6\ mm^4$ Assumed.

① Deflection under each load.

② Maximum Deflection.

Given: $I = 85\times10^5\ mm^4$; $E = 2\times10^5\ N/mm^2$

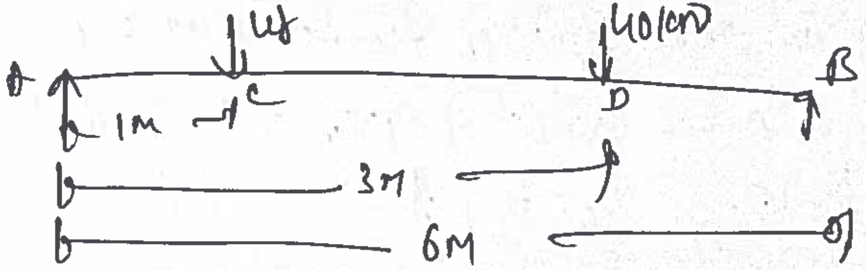
First calculate the reactions RA and RB.

Taken moment about A, we get

$$R_A \times 6 = 48\times1 + 40\times3 = 168$$

$$R_B = 168/6 = 28kN.$$

$$R_A = \text{Total load} - R_B = (48+40)-28 = 60kN$$



Consider the section x in the last part of the beam (i.e. in length DB) at a distance x from the left support A. The BM at this section is given by,

$$EI \frac{d^2y}{dx^2} = R_A x \;\Big|\; -48(x-1) \;\Big|\; -40(x-3)$$

$$= 60x \; \vdots \; - 48(x-1) \; \vdots \; - 40(x-3)$$

Integrate the above equation, we get

$$EI\frac{dy}{dx} = \frac{60x^2}{2} + C_1 \; \vdots \; - \frac{48(x-1)^2}{2} \; \vdots \; - \frac{40(x-3)^2}{2}$$

$$= 30x^2 + C_1 \; \vdots \; - 24(x-1)^2 \; \vdots \; - 20(x-3)^2$$

Integrate the above equation, again we get

$$EIy = \frac{30x^3}{3} + C_1x + C_2 \; \vdots \; - \frac{24(x-1)^3}{3} \; \vdots \; - \frac{20(x-3)^3}{3}$$

$$= 10x^3 + C_1x + C_2 \; \vdots \; - 8(x-1)^3 \; \vdots \; - \frac{20}{3}(x-3)^3 \quad \text{—(iii)}$$

To find value of C_1 and C_2, use two boundary conditions.

(i) at $x = 0$; $y = 0$ and (ii) at $x = 6$ m; $y = 0$.

(i) Substitute of the first boundary condition i.e at $x = 0$; $y = 0$ in equation (iii) and consider the equation up to first dotted line,

$$0 = 0 + 0 + C_2 \; ; \; C_2 = 0$$

(ii) Substitute the second boundary condition i.e at $x = 6$ m, $y = 0$ in equation (iii) and consider the complete equation as $x = 6$ lies in the last part of the beam. we get

$$0 = 10 \times 6^3 + C_1 \times 6 + 0 - 8(6-1)^3 - \frac{20}{3}(6-3)^3$$

$(C_2 = 0)$

$$0 = 2160 + 6C_1 - 8 \times 5^3 - \frac{20}{3} \times 3^3$$

$$= 2160 + 6C_1 - 1000 - 180 = 980 + 6C_1$$

$$C_1 = -980/6 = -163.33$$

Now substitute the value of C_1 and C_2 in equation (iii) we get

$$EIy = 10x^3 - 163.33x \; \vdots \; - 8(x-1)^3 \; \vdots \; - \frac{20}{3}(x-3)^3 \quad \text{—(iv)}$$

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ANSWER KEY AND SCHEME OF EVALUATION

(i) Deflection under first load! at point C; $x=1$ equation (iii) up to first dotted line (the point C lies in the first part of the beam)

$$EI\,y_C = 10\times1^3 - 163.33\times1$$

$$= 10 - 163.33 = -153.33\,kN\,m^3$$

$$= -153.33\times10^3\,Nm^3$$

$$= -153.33\times10^3\times10^6\,Nmm^3$$

$$= -153.33\times10^{12}\,Nmm^3$$

$$y_C = \frac{-153.33\times10^{12}}{2\times10^5\times85\times10^6}\,mm$$

(ii) Maximum Deflection! The deflection is likely to be maximum at section b/w C and D. For maximum deflection, dy/dx should be zero

$$30x^2 + C_1 - 24(x-1)^2 = 0$$

$$30x^2 - 163.33 - 24(x^2+1-2x) = 0$$

$$6x^2 + 48x - 187.33 = 0$$

$$x = \frac{-48 \pm \sqrt{48^2+4\times6\times187.33}}{2\times6} = 2.87m.$$

$$EI\,y_{max} = 10\times2.87^3 - 163.33\times2.87 - 8(2.87-1)^3$$

$$= 236.39 - 468.75 - 52.3$$

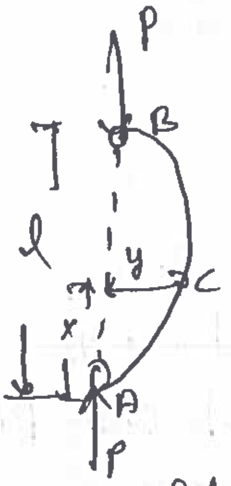
$$= 284.67\,kN\,m^3$$

$$= -284.67\times10^{12}\,Nmm^3$$

$$y_{max} = \frac{-284.67\times10^{12}}{2\times10^5\times85\times10^6} = -16.74mm$$

12) Derive an expression for crippling load when both ends of the column are hinged?

Ans:



The load at which the column just buckles (or bends) is called crippling load.

- Consider a column AB of length l and uniform cross section area hinged at Both of its ends A and B.
- Let P be the crippling load at which the column has just buckled.
- Due to crippling load, the column will deflect into a curved form ACB.
- Consider any section at a distance x from the end A.

Let y = deflection (lateral displacement) at the section.

The moment due to crippling load at the section = $-Py$

(-ve sign taken due to sign convention)

$$M = EI\frac{d^2y}{dx^2}$$

Equating two moments we have, $EI\frac{d^2y}{dx^2} = -Py$ or

$$EI\frac{d^2y}{dx^2} + Py = 0$$

$$\frac{d^2y}{dx^2} + \frac{P}{EI} \cdot y = 0$$

The solution of the above differential equation

$$y = C_1 \cos\left(x\sqrt{P/EI}\right) + C_2 \sin\left(x\sqrt{P/EI}\right)$$

ANSWER KEY AND SCHEME OF EVALUATION

When C_1 and C_2 are the constants of Integration, the values of C_1 and C_2 are obtained as given below.

① At A, $x=0$ and $y=0$

Substituting these value in equation ①, we get

$$0 = C_1 \cos 0 + C_2 \sin 0$$
$$= C_1 \times 1 + C_2 \times 0$$
$$= C_1$$
$$\boxed{C_1 = 0} \quad \text{— (1a)}$$

② At B, $x = l$, $y = 0$,

$$0 = C_1 \cos\left(l \times \sqrt{P/EI}\right) + C_2 \sin\left(l \times \sqrt{P/EI}\right)$$
$$= 0 + C_2 \sin\left(l \times \sqrt{P/EI}\right)$$
$$= C_2 \sin\left(l \sqrt{P/EI}\right) \quad \text{— (2a)}$$

From equation (2a), it is clear that $C_2 = 0$

$$\sin\left(l \sqrt{P/EI}\right) = 0$$

At $C_1 = 0$; then If C_2 is also equal to zero ① we will get

$y = 0$.

$$\sin\left(l \sqrt{P/EI}\right) = 0$$
$$= \sin 0 \text{ or } \sin \pi \text{ or } \sin 2\pi \text{ or } \sin 3\pi \text{ or } \ldots$$

or

$$l \sqrt{P/EI} = 0 \text{ or } \pi \text{ or } 2\pi \text{ or } 3\pi$$

Taken least practical value

$$l \sqrt{P/EI} = \pi$$

$$\boxed{P = \frac{\pi^2 EI}{l^2}}$$

⑱

13) A 1.75m long steel column of rectangular cross section 120mm x 100mm is rigidly fixed at one end and hinged at the other. Determine the buckling load on the column and the corresponding axial stress using Euler's formula. Take E for the column material as 200GPA.

<u>Ans:</u> <u>Given data:</u> length of steel column = 1.75m.

rectangular c/s = 120 x 100mm

End conditions = ~~Both~~ one end is fixed and other end is hinged

$$l_e = l/\sqrt{2}$$

$$L = 1.75M \rightarrow l_e = \frac{1.75}{\sqrt{2}} = 1.237M$$

$$d = 120\,m \quad b = 100\,m$$

$$P_d = \frac{\pi^2 EI}{l^2} = \frac{\pi^2 \times (200 \times 10^3) \times \left(\frac{100 \times 120^3}{12}\right)}{(1237)^2}$$

$$= 18576.02 \times 10^3 N$$

buckling load = 18576.02 kN

4) Axial stress on column = $\frac{\text{load}}{\text{Area}}$

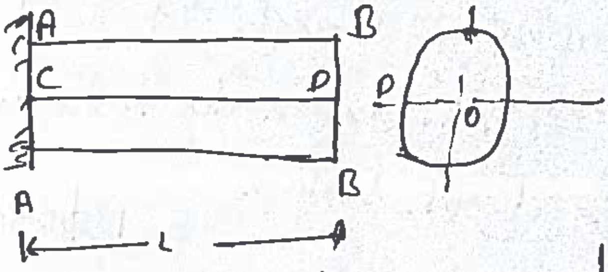
$$= \frac{18576.02}{120 \times 100} \times 1000$$

$$= 1548.00 N/mm^2$$

ANSWER KEY AND SCHEME OF EVALUATION

14) Derive torsion equation?

Ans:



- when a circular shaft is subjected to torsion, shear stresses are setup in the material of the shaft.
- To determine the magnitude of shear stress at any point on the shaft, consider a shaft fixed at one end AA and free at the end BB as shown in fig.
- let CD is any line on the outer surface of the shaft.
- The shaft is subjected to torque at the end BB as shown fig
- As a result of this torque T, the shaft at the end BB will rotate clockwise and every crossection of the shaft will be subjected to shear stress.
- The point D will shift to D' and hence line CD will be deflected to CD' as shown fig.
- The line OD will be shifted to OD'

let R = Radius of shaft

L = length of shaft

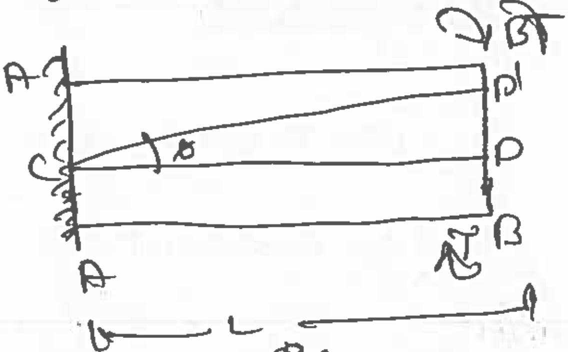
T = Torque applied at the end BB

τ = Shear stress induced at the surface of the shaft due to torque T

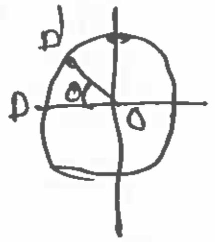
C = Modulus of rigidity of the material of the shaft.

ϕ = $\angle DCD'$ also called the shear strain

θ = $\angle DOD'$ and is also called angle of twist



(a)



(b)

Shaft fixed at AA and subjected to torque T at BB

Now distortion at the outer surface due to torque T = DD'

Shear strain at outer surface = Distortion per unit length

$$= \frac{\text{Distortion at the outer surface}}{\text{length of shaft}} = \frac{DD'}{L}$$

$$= \frac{DD'}{CD} = \tan\phi$$

$= \phi$ (If ϕ is very small then $\tan\phi = \phi$)

∴ Shear strain at the outer surface

$$\phi = \frac{DD'}{L}.$$

Now from fig (b)

$DD' = OD \times \theta = R\theta$ ($OD = R$ = Radius of shaft)

Substitute the value of DD' in equation, we get

Shear strain at the outer surface $\phi = \frac{R \times \theta}{L}$

Now, modulus of rigidity (C) of the material of the shaft

$$C = \frac{\text{Shear stress induced}}{\text{Shear strain produced}} = \frac{\text{Shear stress at outer surface}}{\text{Shear strain at outer surface}}$$

$$= \frac{\tau}{\left(\frac{R\theta}{L}\right)}$$

ANSWER KEY AND SCHEME OF EVALUATION

$$= \frac{\tau}{\left(\frac{R\theta}{L}\right)}$$

$$= \frac{\tau \times L}{R\theta}$$

$$= \frac{C\theta}{L} = \frac{\tau}{R}$$

$$\boxed{\tau = \frac{R \times C \times \theta}{L}}$$

→ now, given shaft subjected to a given torque (T), the value of C, θ, and L are constent.
shear stress is proportional to the radius R.

$\tau \propto R$ or $\frac{\tau}{R} = \text{constant}$

If q is the shear stress induced at a radius "r" from the centre of the shaft then

$$\frac{\tau}{R} = \frac{q}{r}$$

$$\frac{\tau}{R} = \frac{C\theta}{L} \quad \text{from eqn}$$

$$\boxed{\frac{\tau}{R} = \frac{C\theta}{L} = \frac{q}{r}}$$

→ shear stress is maximum at outer surface and shear stress is zero at the axis of the shaft.

15) The stiffness of a closely coiled helical spring is 1.5 N/mm of compression under a maximum load of 100N. The maximum shearing stress produced in the wire of the spring is 130 N/mm2. The solid length of the spring (when the coils are touching) is given 5cm.

Find ① Diameter of the wire.

(ii) mean diameter of the coils.

(iii) No of coils required.

Take $C = 4.5 \times 10^4$ N/mm2

Ans. stiffness of spring $S = 1.5$ N/mm.

Load on spring = $W = 100$ N.

maximum shear stress $\tau = 130$ N/mm2.

Solid length of spring = 5cm = 50mm.

Modulus of rigidity $C = 4.5 \times 10^4$ N/mm2.

Let d = diameter of wire.

D = mean dia of coil

R = mean Radius of coil $D/2 = R$.

n = number of coils.

$$S = \frac{Cd^4}{64R^3 n} = (8) \quad 1.5 = \frac{4.5 \times 10^4 \times d^4}{64 \times R^3 \times n}.$$

$$d^4 = \frac{1.5 \times 64 \times R^3 \times n}{4.5 \times 10^4}$$

$$d^4 = 0.002133\, R^3 \times n \quad \text{—(i)}$$

$$\tau = \frac{16 W \times R}{\pi d^3} \text{ or } 130 = \frac{16 \times 100 \times R}{\pi d^3}$$

$$R = \frac{130 \times \pi d^3}{16 \times 100} =$$

$$\boxed{R = 0.4090 d^3} \quad \text{—(ii)}$$

substitute the value of R in equat (i), we get

$$d^4 = 0.002133 \times (0.40906 d^3)^3 \times n$$

$$= 0.002133 \times (0.40906)^3 \times d^9 \times n = 0.00014599 \times d^9 \times n$$

$$\frac{d^9 . n}{d^4} = \frac{1}{0.00014599} \;\&\; d^5 . n = \frac{1}{0.00014599}$$

Solid length = n×d or 50 = n×d

$$n = \frac{50}{d}.$$

(½)

substitute the value of n equation (iii)

$$d^5 \times \frac{50}{d} = \frac{1}{0.00014599}$$

$$d^4 = \frac{1}{0.00014599} \times \frac{1}{50} = \underline{136.95}$$

$$n = \frac{50}{d} = \frac{50}{3.42} = 14.62 \approx \underline{15}$$

$$R = 0.40906\,d^3 = 0.40906 \times (3.42)^3$$

$$R = \underline{16.36mm}$$

Mean dia of coil $D = 2R = 2 \times 16.36$

$$= \underline{32.72M}$$

$$\boxed{D = 32.72m}$$

a.p.

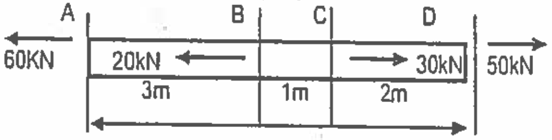
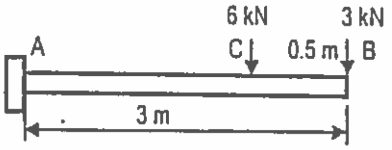
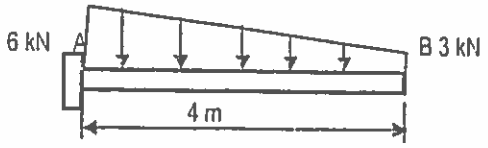
Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	Program	ME			Academic Year	2021 - 2022
Course Code	20ME303	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	**Mechanics of Solids**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Give the relationship between bulk modulus and modulus of elasticity in terms of Poisson's ratio.	20ME303.1	L1
2	Differentiate the point load, UDL and VDL.	20ME303.2	L1
3	Define the flexural rigidity and torsional rigidity.	20ME303.2	L2
4	Mention the advantages of hollow circular shafts over solid circular shafts.	20ME303.2	L1
5	Define buckling and stability.	20ME303.1	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw a neat stress- strain curve diagram of stainless steel and explain	6M	20ME303.2	L2
6 (b)	A steel bar is 900 mm long ,its two ends are 40 mm and 30 mm in dia and the length of each rod is 200 mm The middle portion of the bar is 15 mm in dia and 500 mm long ,if the bar is subjected to an axial tensile load of 15 KN .Find total extension. Take E=200 GPa	6M	20ME303.2	L2
	OR			
7 (a)	A steel bar of 20 mm diameter is acted upon by the forces. What is the elongation of the bar when young's modulus, E = 210 GPa. Find net elongation by principal of super position  	6M	20ME303.3	L2
7 (b)	The stresses on two perpendicular planes through a point in a body are 120 MPa (Tensile) and 80 MPa (Compression). Determine the normal and tangential stress on a plane at an angle 40⁰ with the vertical (ACW). Draw configuration and Mohr's diagrams	6M	20ME303.3	L2
8 (a)	Write about the types of beams in detail with neat sketches	6M	20ME303.2	L2
8 (b)	Draw shear force and bending moment diagrams for a cantilever beam and find shear force and bending moments of span carrying two point loads 3 kN and 6 kN at right end and 0.5 m from right end  	6M	20ME303.2	L2
	OR			
9 (a)	Differentiate the shear force and bending moments when point, uniformly distributed and variably distributed loads applied.	6M	20ME303.2	L2
9 (b)	A cantilever beam 4 m long carries a VD, 2 kN/m at the free end to 5 kN/m at the fixed end and draw SFD and BMD  	6M	20ME303.2	L2
10 (a)	Write the sign convention of shear force and bending moment to the	6M	20ME303.2	L2

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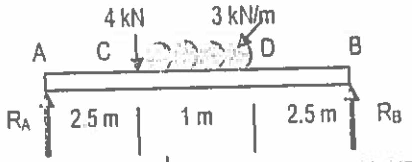
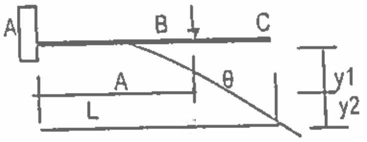
	left of the section.			
10 (b)	Draw the free body diagram, shear force and bending moment diagram of VDL on simply supported beam	6M	20ME303.3	L2
	OR			
11 (a)	The simply supported beam is 6 m carries a point load 4 kN at a distance of 2.5 m from left where the UDL of 3 kN/m starts from point load for 1 m. Find the reactions and draw the SFD and BMD  	6M	20ME303.3	L4
11 (b)	Derive the equations for simply supported beam with UDL an UVL	6M	20ME303.2	L2
12 (a)	Derive equation for moment of inertia for a I- section	5M	20ME303.2	L2
12 (b)	A soli circular shaft transmits a power of 60 KW at 200 rpm. Find the diameter of the shaft if the allowable shear stress is 50 MPa and allowable twist is 2⁰ for every 10 mt length of shaft. Take C=80 GPa.	7M	20ME303.2	L2
	OR			
13 (a)	A cantilever beam of span L is subjected to a UDL of W KN/m at a distance 'a' from fixed end. Find the deflection of free end  	5M	20ME303.2	L2
13 (b)	Explain the Macaulay's method in deflection of beams	7M	20ME303.3	L2
14 (a)	Explain the buckling failure in columns with Rankine formula	6M	20ME303.2	L3
14 (b)	Derive an expression for Euler's critical load with one fixed an the other end hinged.	6M	20ME303.2	L2
	OR			
15 (a)	Distinguish between circumferential and longitudinal stress in cylindrical shell when subjected to an internal pressure.	5M	20ME303.2	L2
15 (b)	Differentiate thin and thick cylinders and write three applications of compound cylinders.	7M	20ME303.2	L2

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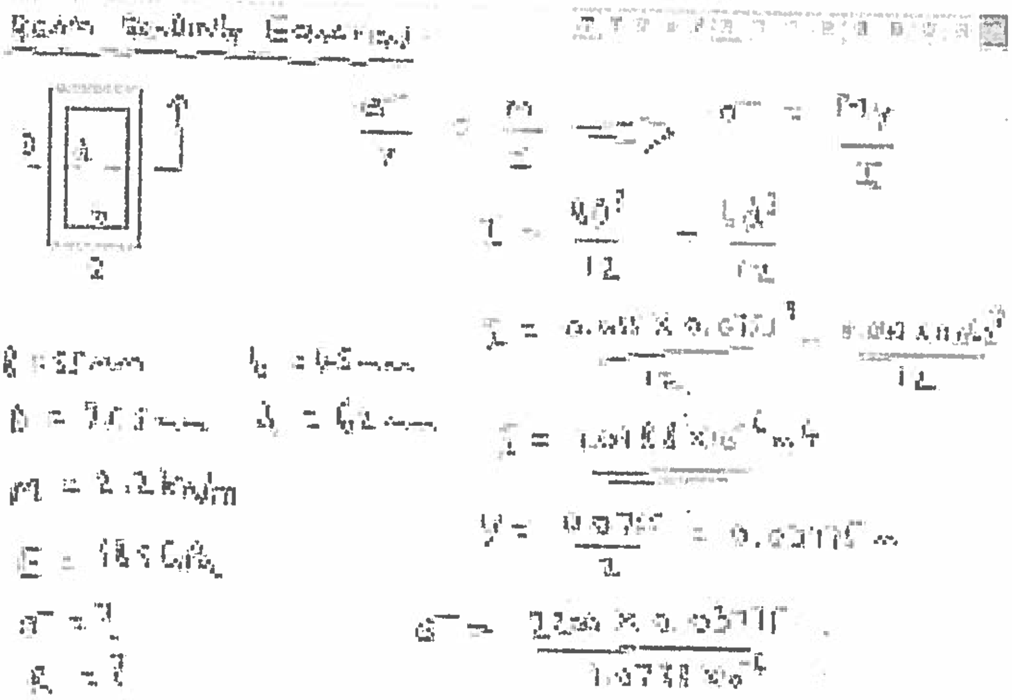
Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	Program	ME			Academic Year	2021 - 2022
Course Code	20ME303	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	MECHANICS OF SOLIDS						

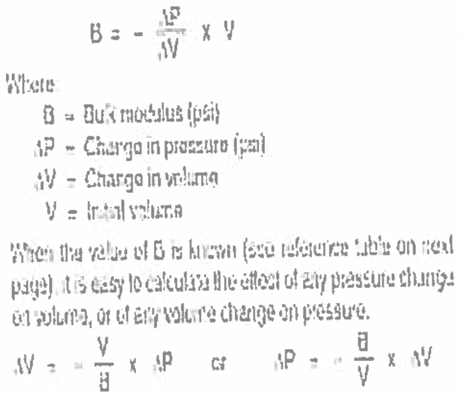
No.	Questions (1 through 5)
1	Define the bulk modulus and modulus of elesticity. To put in more simple words, the bulk modulus is nothing but a numerical constant that is used to measure and describe the elastic properties of a solid or fluid when pressure is applied on all the surfaces. The bulk modulus of elasticity is one of the measures of the mechanical properties of solids
2	Differentiate the point load, UDL and VDL. When placed in steel storage racks, a uniformly distributed load is one whose weight is evenly distributed over the entire surface of the rack's beams or deck. A point load is a one with its weight significantly concentrated in one (or more) places on the rack's beams or decks.
3	M/I=f/y=E/R – justify. Bending Equation. The axial deformation of the beam due to external load that is applied perpendicularly to a longitudinal axis is called the Bending Theory. The bending equation stands as f/y = E/R = M/I. 

Explain pure torsion

4 When the circular shaft is subjected to torque only without being acted upon by any bending moment or axial force, the circular shaft is said to be in the state of pure torsion.

Define buckling and stabilitythe state of being stable

5



$$B = -\frac{\Delta P}{\Delta V} \times V$$

Where:

B = Bulk modulus (psi)

ΔP = Change in pressure (psi)

ΔV = Change in volume

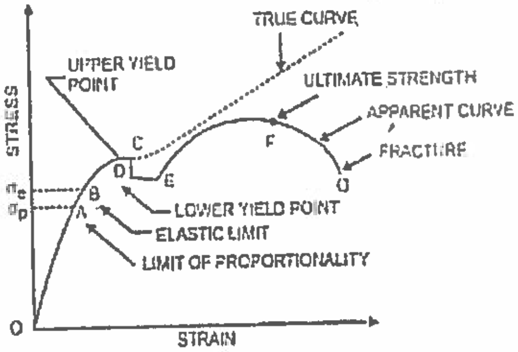
V = Initial volume

When the value of B is known (see reference table on next page), it is easy to calculate the effect of any pressure change on volume, or of any volume change on pressure.

$$\Delta V = -\frac{V}{B} \times \Delta P \quad \text{or} \quad \Delta P = -\frac{B}{V} \times \Delta V$$

No.	Questions (6 through 15)	Marks
	Draw a neat stress- strain curve diagram of stainless steel and explain.	

6 (a)



6M

6 (b) A steel bar of 900 mm long, its ends are 40mm and 30mm in dia. and the length of each rod is 200mm. The middle portion of the bar is 15mm in dia. If the cylinder is carrying a load of 20kN, find the total extension when modulus of elasticity is 200 Gpa. 6M

. Area of steel bar =2800mm2

Data: Area of steel, As = 16 cm2
Area of brass, Ab = 10 cm2
Weight, W = 500 kg

Δl is same for steel and brass, so

2m

As=yl δl

The sum of forces of steel and brass is, fs+2fb=5000 -------(1)

Ss=(fs)/ (as) , sb=(fb)/ (ab)

Fs=ss.as, fb=sb.ab

Ss=(ys)(is)(δl)

Ss is directly proportional to isys

Similarly, sb is directly proportional to ibyb

Where is and ib are the moment of inertia of steel and brass

2m

s is the stress

Sbss=ybisyslb=2×32=34

Ss=34sb

Ssas+2(ssab)=5000 -----from eq. (1)

(ss)(as+2×43)(ss.ab)=5000

Ss (16+23×10)=5000

Ss(31)=5000

Ss=315000=161kg/cm2

Sb=43×115000

Sb=121kg/cm2

A steel bar of 20 mm diameter is acted upon by the forces. What is the elongation of the bar when young's modulus, E = 210 GPa. Find net elongation by principal of super position.

Sol: Length, l = 5000 mm; cross sectional area, A= 22/7x20x20/4 =400mm2;

7 (a) Tensile force, P =50 KN = 50 x 103 N; Young's Modulus, E = 200 x 106 N/mm2. 6M

2M

As per the young's modulus, E = [(P/A) / (δl / l)]

Therefore, change in length or elongation,

δl = (Pl / AE) 2M

δl = (50 x 103 x 103)/ (400 x 200 x 103)

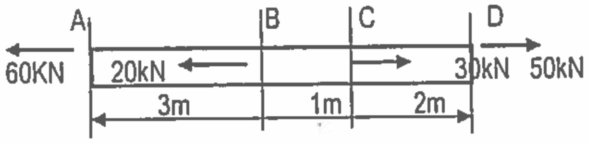
= ½ = 0.6 mm

2M

7A. Area of steel bar =2800mm2

Data: Area of steel, As = 16 cm2
Area of brass, Ab = 10 cm2
Weight, W = 500 kg

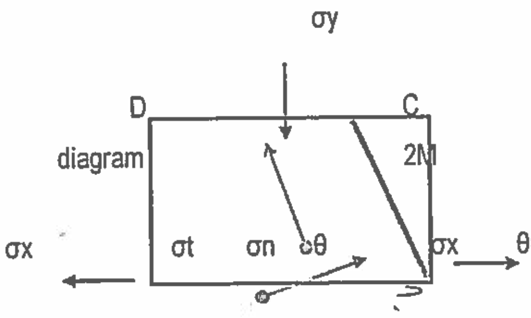
Δl is same for steel and brass, so



The stresses on two perpendicular planes through a point in a body are 120 MPa(Tensile) and 80 MPa(Compression). Determine the normal and tangential stress on a plane at an angle 300 with the vertical (ACW).Draw configuration and Mohr's diagrams.

The stresses on two perpendicular planes through a point in a body are 90 MPa (Tensile) and 70 MPa(Compression). Determine the normal and tangential stress on a plane at an angle 250 with the vertical (ACW).

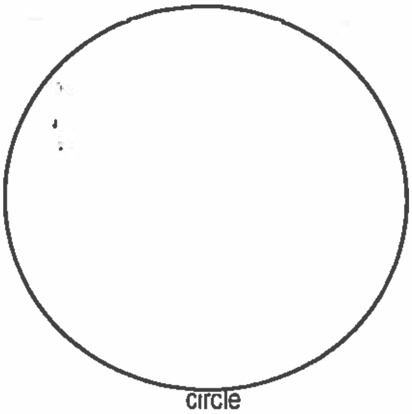
7 (b) 6M



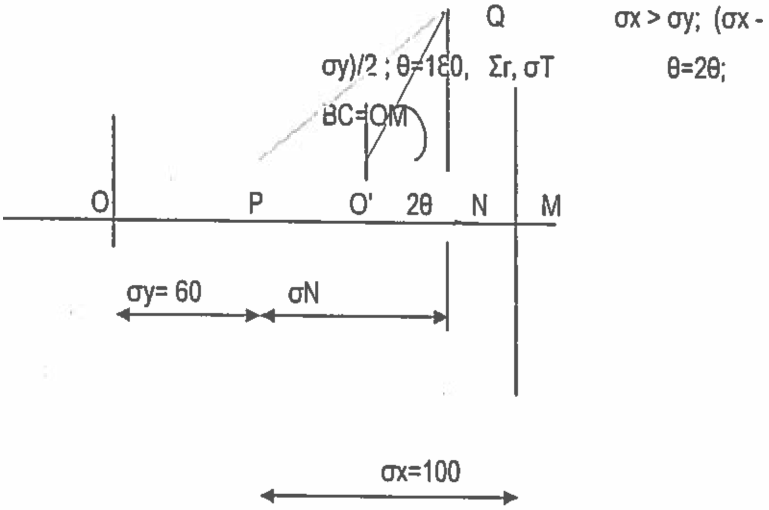
Configuration

θ=θacw; angle b/w σx & σy = 90

Mohr's



circle



σN = ?; σT=?; σR=   σR = √σN2 + σT2 =

4M

6B.

A steel rod of 5 m long and 20mm diameter is subjected to load 50 kN. Find the elongation when young's modulus is 2 x 106 N/mm2. Poison's ratio is 0.25

Sol: Length, l = 5000 mm; cross sectional area, A= 22/7x20x20/4 =400mm2; Tensile force, P =50 KN = 50 x 103 N; Young's Modulus, E = 200 x 106 N/mm2.

2M

As per the young's modulus, E = [(P/A) / (δl / l)]

Therefore, change in length or elongation, δl = (Pl / AE)

2M

δl = (50 x 103 x 103)/ (400 x 200 x 103)

= ½ = 0.6 mm

8 (a) Write about the types of beams.

Simply supported beam

Both end hinged beam

Fixed beam

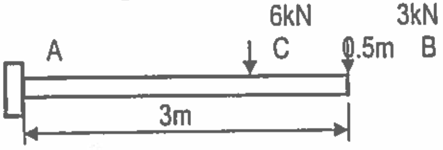
Continuous beam

Cantilever beam

Propped cantilever beam

6M

8 (b) Draw shear force and bending moment diagrams for a cantilever beam and find shear force and bending moments of span carrying two point loads 2 kN and 5kN at right end and 0.5m from right end.

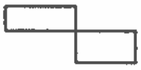


6M

9 (a) Differentiate the shear fo

rce and bending moments when point, uniformly distributed and variably distributed

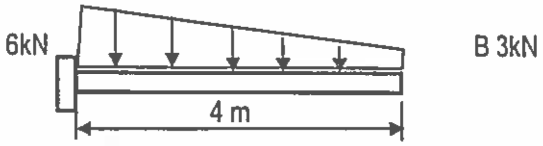
6M

Load	Slope for shear force	Slope for bending Moment
P 	Constant 	Linear 
Uniformly distributed load 	Linear 	Parabolic 
Uniformly varying load 	Parabolic 	Cubic 

loads applied. Figure-1 Slopes for various types of loads

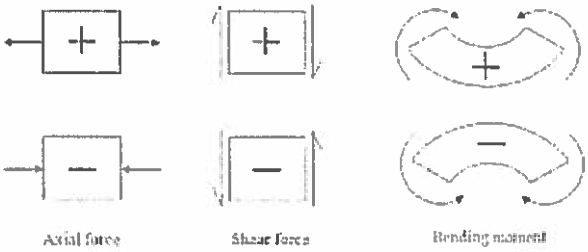
A cantilever beam 4m long carries a VD, 2kN/m at the free end to 5kN/m at the fixed end and draw SFD and BMD.

9 (b) 6M



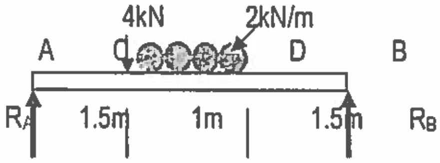
Write the sign convention of shear force and bending moments.

10 (a) 6M

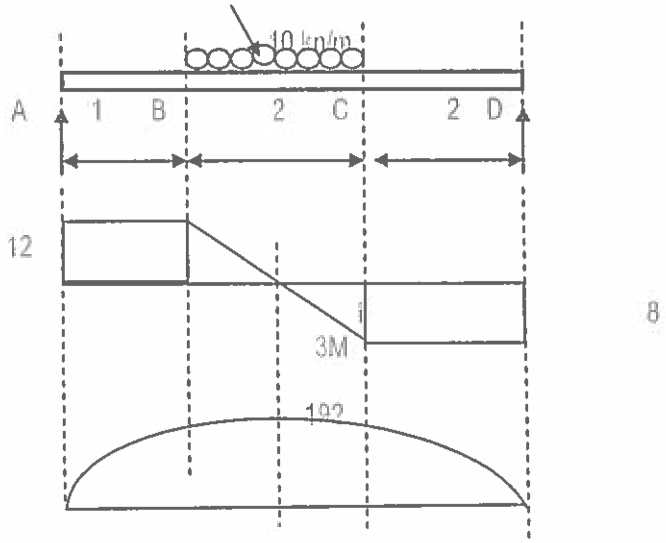


10 (b) Draw the free body diagram, shear force and bending moment diagram of VDL. 6M

The simply supported beam is 5 m carries a point load 4 kN at a distance of 1.5m from left where the UDL of 2kN/m strarts from point load for 1 m. Find the reactions and draw the SFD and BMD.



11 (a) 6M



Ma =0

Mc =12x 1=12 kn-m

Md=0 =8x 2=16 kn-m

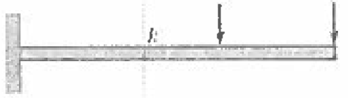
x/12 =(2-x)/8

x= 1.2 kn-m

Mm = 12(1+1.2)-10x1.2xq.2/2 = 19.2 kn-m

3M

Derive the equations for simply supported beam with UVL.



(a)

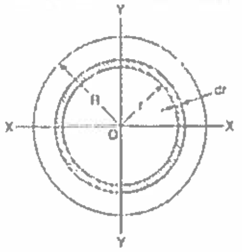
11 (b) 6M

(b)

12 (a) Derive equation for moment of inertia for a rect angular section. 5M

Derive an equation for Tortion.

12 (b) 7M



Total torque could be easily determined by integrating the above equation between limits 0 and R.

Therefore, total torque transmitted by a circular solid shaft is , $T= t/R\int r^2dA$ -------------------- 1 4M

Polar moment of inertia is ------------------------- 2

$J=\int r2\, dA$

Therefore total torque transmitted by a circular solid shaft is,

$T=tJ/R$

$T/J= t/R$

4M

The expression for shear stress produced in a circular shaft subjected to torsion and therefore we have following result from that expression

$T/J = C\theta/L$

Considering above two equations, the torsion equation for circular shaft is, $T/J = t/R = C\theta/L$ --------3m 4M

Where,

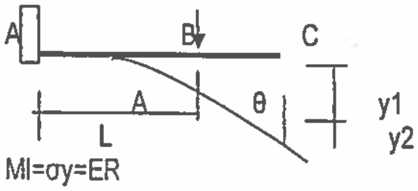
C = Modulus of rigidity

L = Length of shaft

θ = Angle of twist

A cantilever beam of span L is subjected to a concentrated load W at a distance 'a' from fixed end. Find the deflection of free end.

13 (a)



5M

$MI=\sigma y=ER$

Where, E = Modulus of Elasticity of beam material

R = Radius of Curvature of the beam

M = Moment of resistance

I = Moment of Inertia of beam c/s about N.A

σσ = Bending stress at the layer situated at a distance y from N.
3M

y = Distance of layer from N.A.

$\sigma/y = E/R$ -------1

At the distance 'y', let us consider an elementary strip of very small thickness dy.

'σ' is the bending stress in this strip.

Let dA = area of this elementary strip.

The force developed in this strip = σ.dA.

The elementary moment of resistance due to this elementary force is given by dM = f.dA.y 3M

Total moment of resistance due to all such elementary forces is given by

$\int dM = \int \sigma \times dA \times y$ or $M = \int \sigma \times dA \times y$ ----- 2

From Eq. 1, $\sigma = y \times (E/R)$ Putting this value of f in Eq. 2

we get where I = Moment of inertia of the whole area about the neutral axis N-A.

Where; M = Bending moment I = Moment of Inertia about the axis of bending i.e; Ixx

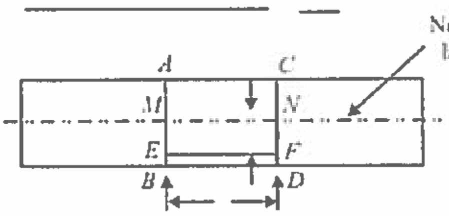
y = Distance of the layer

The extreme distance of extreme fibre from N.A.)

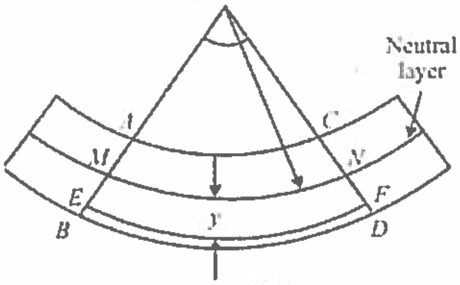
E = Modulus of elasticity of the beam material.

R = Radius of curvature

3M



Before bending



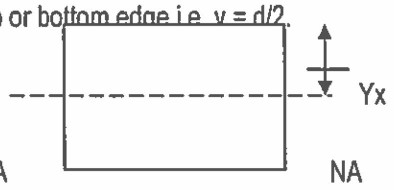
After bending

Explain the Macaulay's method in deflection of beams.

13 (b) The modulus of section may be defined as the ratio of moment of inertia to the distance to the extreme fibre. It is denoted by Z. Z= I/y ; For rectangular 7M

section, $I = bd^3/12$ & $y = d/2$. $Z= bd^2/6$.

It is the ratio of moment of inertia (I) of the beam cross-section about the neutral axis to the distance (y_{max}) of extreme fiber from the neutral axis. y = distance from the centroid to top or bottom edge i.e. $y = d/2$. 2M



2M

It is the ratio of moment of inertia (I) of the beam cross-section about the neutral axis to the distance (y_{max}) of extreme fiber from the neutral axis. y = distance from the centroid to top or bottom edge i.e. y = d

14 (a) Explain the buckling. 6M

Derive Euler's formula.

$MI=\sigma y=ER$

Where, E = Modulus of Elasticity of beam material

R = Radius of Curvature of the beam

M = Moment of resistance

I = Moment of Inertia of beam c/s about N.A

$\sigma\sigma$ = Bending stress at the layer situated at a distance y from N.

3M

y = Distance of layer from N.A.

$\sigma/y = E/R$ ------1

14 (b) At the distance 'y', let us consider an elementary strip of very small thickness dy. 6M

'σ' is the bending stress in this strip.

Let dA = area of this elementary strip.

The force developed in this strip = σ.dA.

The elementary moment of resistance due to this elementary force is given by dM = f.dA.y 3M

Total moment of resistance due to all such elementary forces is given by

$\int dM = \int \sigma \times dA \times y$ or $M = \int \sigma \times dA \times y$ ---- 2

From Eq. 1, $\sigma = y \times (E/R)$ Putting this value of f in Eq. 2

we get where I = Moment of inertia of the whole area about the neutral axis N-A.

Where; M = Bending moment I = Moment of Inertia about the axis of bending i.e; Ixx

y = Distance of the layer

The extreme distance of extreme fibre from N.A.)

E = Modulus of elasticity of the beam material.

R = Radius of curvature

15 (a) What is circumferential and longitudinal stresses Define circumferential stresses, f1 5M

The hoop stress is the force over area exerted circumferentially (perpendicular to the axis and the radius of the object) in both directions on every particle in the cylinder wall.

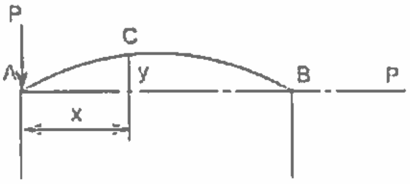
3M

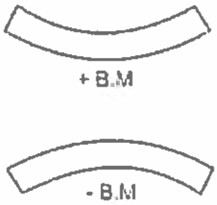
Define longitudinal stresses, f2

Longitudinal stress is defined as the stress produced when a pipe is subjected to internal pressure.

Longitudinal stress is also known as axial stress..

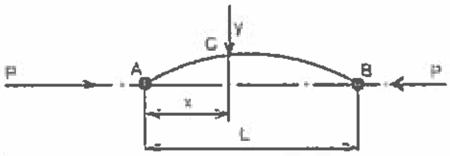
Differentiate thin and thick cylinders.





According to sign convention

4M

15 (b)  7M

4M

Consider an axially loaded strut, shown below, and is subjected to an axial load 'P' this load 'P' produces a deflection 'y' at a distance 'x' from one end.

Assume that the ends are either pin jointed or rounded so that there is no moment at either end.

$B\sin\{L\sqrt{p/EI}\} = 0$

$B=0$ or $\sin\{L\sqrt{p/EI}\} = 0$

$\sin\{L\sqrt{p/EI}\} = 0$ or $\{L\sqrt{p/EI}\} = \pi$ or $nL = \pi$

4M

or $\{\sqrt{p/EI}\} = \pi/L$ or $P = \pi^2EI/L^2$

NSRIT

Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	Program	EEE			Academic Year	2021 - 2022
Course Code	20EE304	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	DC MACHINES & TRANSFORMERS						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Describe multiply excited magnetic field system.	20EE304.1	L1
2	What is armature reaction? What are the effects of armature reaction on the performance of dc machine?	20EE304.2	L1
3	Why a starter is required to start a DC motor? What is the essential element of a starter?	20EE304.3	L2
4	What is the function of a transformer?	20EE304.4	L1
5	Define Sumpner's test.	20EE304.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Derive the expression for field energy and co energy in a doubly excited system assuming constant current system.	6M	20EE304.1	L2
6 (b)	Derive the expression for torque in a singly excited system.	6M	20EE304.1	L2
	OR			
7 (a)	For a singly excited system derive the expression for magnetic field energy stored.	6M	20EE304.1	L2
7 (b)	Write in brief about multiple-excited magnetic field system.	6M	20EE304.1	L2
8 (a)	Explain the constructional details of DC generator.	7M	20EE304.2	L2
8 (b)	Calculate the emf generated by a 4 pole wave wound armature having 45 slots with 18 conductors per slot. When driven at 1200 rpm the flux per pole is 0.016Wb.	5M	20EE304.1	L3
	OR			
9 (a)	Explain magnetization characteristics of a DC shunt generator?	6M	20EE304.2	L2
9 (b)	Express the EMF equation of a DC generator.	6M	20EE304.2	L2
10 (a)	Derive the torque equation of a DC motor.	6M	20EE304.3	L2
10 (b)	A dynamo has a rated armature current at 250A. What is the current per path of the armature if the armature winding is lap or wave connected? The machine has 12 poles.	6M	20EE304.3	L3
	OR			
11 (a)	Explain speed control of DC shunt motor.	6M	20EE304.3	L2
11 (b)	Explain in detail the losses in DC motor.	6M	20EE304.3	L2
12 (a)	Derive the EMF equation of transformer.	6M	20EE304.4	L2
12 (b)	A 400/230V, 50Hz, single phase transformer has 200 turns on high voltage side. Find turns ratio, transformation ratio, and number of turns on low voltage winding. Also find the flux developed in the core.	6M	20EE304.4	L3
	OR			
13 (a)	Draw the various types of three phase transformer connections and brief each one of them.	6M	20EE304.4	L3
13 (b)	Draw and explain the phasor diagram of single phase transformer on load considering with winding resistance.	6M	20EE304.4	L2
14 (a)	Draw the circuit diagram of Sumpner's test and derive the equation for efficiency of each transformer.	6M	20EE304.5	L2
14 (b)	The primary and secondary voltages of an autotransformer are 1200V and 600V respectively. Calculate the economy of Cu when the secondary current is 120A. Draw the circuit and show the current distribution in the winding.	6M	20EE304.5	L2
	OR			

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15 (a)	A 10 kVA, 500/250 V, 50 Hz single-phase transformer gave the following test data: OC Test (LV side): 250 V, 1.0 A, 80 W SC Test (HV side): 25 V, 12 A, 100 W Where LV refers to the low voltage and HV refers to high voltage side. Determine the following: (i) Equivalent circuit referred to LV side (ii) Secondary load voltage at 0.8 p.f. lagging with full-load current.	7M	20EE304.5	L3
15 (b)	Explain how the Scott connection can be used to obtain two phase supply from a three phase supply.	5M	20EE304.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

DCMT SUPPLY EXAMINATION KEY & SCHEME

PART A

1) **Describe multiply excited magnetic field system?**

Ans: Electro-mechanical transducers have the special requirement of producing an electrical signal proportional to forces or velocities or producing force proportional to electrical signal (current or voltage). Such transducers require two excitation one excitation establishes a magnetic field of specified strength while the other excitation produces the desired signal (electrical or mechanical).Also continuous energy conversion devices motors and generators require multiple excitation.

2) **what is armature reaction ? what are the effects of armature reaction on the performance of dc machines?**

Ans: **The current flowing through the armature conductors creates a magnetic field, which is called as armature flux.** This armature flux distorts and weakens the magnetic flux produced by the main poles. This effect of armature flux on the main flux is known as armature reaction.

3) **why a starter is required to start a dc motor? What is the essential element of a starter?**

Ans) Starters are used to protect DC motors from damage that can be caused by very high current and torque during startup. They do this by providing external resistance to the motor, which is connected in series to the motor's armature winding and restricts the current to an acceptable level.

4) **What is the function of a transformer?**

Ans) A transformer is a device that **transfers electric energy from one alternating-current circuit to one or more other circuits, either increasing (stepping up) or reducing (stepping down) the voltage.**

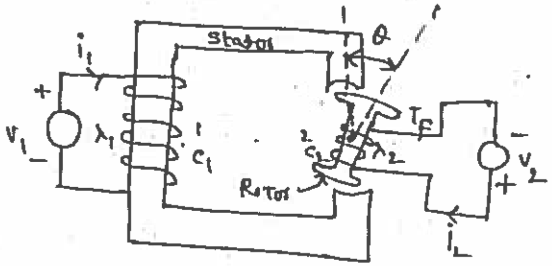
5) **Define sumpners test?**

Ans) **A full-load test on the large transformer is to be conducted to determine the maximum temperature rise of the transformer** which is called as back-to-back test. The back-to-back test is also known as Sumpner's test or regenerative test.

PART B

6a) **Derive an expression for field and co energy in a doubly excited system assuming constant current system?**

Multiply – excited Magnetic field systems



→ single exited devices are generally employed for motion through a limited distance or rotation through a prescribed angle.

→ Electro mechanical Transducers have the special requirement of producing an electrical signal proportional to forces or velocities or producing force proportional to electrical signal (current or voltage)

→ such Transducers require two excitations one excitation establishes a magnetic field while the other excitation produces the desired signal (electrical or mechanical)

→ figure shows a magnetic field system with Two electrical excitations – one on stator and the other on rotor

→ The System can be described in either of the two sets of three independent variables $(\lambda_1 \lambda_2\ \theta)$ or $(i_1\ i_2\ \theta)$

Ans)

In terms of the first set

$$T_f = \frac{\partial W_f}{\partial \theta}(\lambda_1, \lambda_2, \theta) \quad —①$$

Where the field energy is given by

$$W_f(\lambda_1, \lambda_2, \theta) = \int_0^{\lambda_1} i_1 \, d\lambda_1 + \int_0^{\lambda_2} i_2 \, d\lambda_2 \quad —②$$

we know that $i = \frac{\partial W_f(\lambda, x)}{\partial \lambda}$ Like this

$$i_1 = \frac{\partial W_f}{\partial \lambda_1}(\lambda_1, \lambda_2, \theta)$$

$$i_2 = \frac{\partial W_f}{\partial \lambda_2}(\lambda_1, \lambda_2, \theta)$$

Assuming linearity

$$\lambda_1 = L_{11} i_1 + L_{12} i_2$$

$$\lambda_2 = L_{21} i_1 + L_{22} i_2$$

$$(L_{12} = L_{21})$$

∵ The flux produced by C_1 links with its own coil C_1 & secondary coil C_2

$$\Phi_1 = L_1 i_1 + M_1 i_2$$

Similarly $\Phi_2 = L_2 i_2 + M_2 i_1$

solving for i_1 & i_2 in terms of λ_1, λ_2 and substituting in eqn ②

$$W_f(\lambda_1, \lambda_2, \theta) = \frac{1}{2}\beta_{11}\lambda_1^2 + \beta_{12}\lambda_1\lambda_2 + \frac{1}{2}\beta_{22}\lambda_2^2$$

where $\beta_{11} = \frac{L_{22}}{L_{11}L_{22} - L_{12}^2}$ $\beta_{12} = \beta_{21} = \frac{-L_{12}}{(L_{11}L_{22} - L_{12}^2)}$

$$\beta_{22} = \frac{L_{11}}{L_{11}L_{22} - L_{12}^2}$$

The self & mutual inductance of two exciting coils are function of angle θ

If currents are used to describe the system state

$$T_f = \frac{\partial W_f'}{\partial \theta}(i_1, i_2, \theta)$$

Coenergy $W_f'(i_1, i_2, \theta) = \int_0^{i_1} \lambda_1 \, di_1 + \int_0^{i_2} \lambda_2 \, di_2$

in linear case $W_f'(i_1, i_2, \theta) = \frac{1}{2}L_{11}i_1^2 + L_{12}i_1 i_2 + \frac{1}{2}L_{22}i_2^2$

where inductances are functions of angle θ

∵ $i_1 = \beta_{11}\lambda_{11} + \beta_{12}\lambda_2$

$i_2 = \beta_{21}\lambda_1 + \beta_{22}\lambda_2$; $\beta_{21} = \beta_{12}$

$$W_f(\lambda_1, \lambda_2, \theta) = \int_0^{\lambda}(\beta_{11}\lambda_1 + \beta_{12}\lambda_2)\, d\lambda_1 + \int_0^{\lambda_2}(\beta_{12}\lambda_1 + \beta_{22}\lambda_2)\, d\lambda_2$$

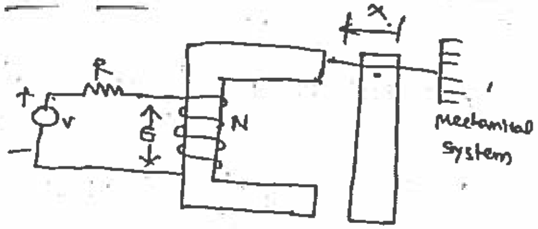
$$= \beta_{11}\int_0^{\lambda_1}\lambda_1 \, d\lambda_1 + \beta_{12}\left[\int_0^{\lambda_1}\lambda_2 \, d\lambda_1 + \int_0^{\lambda_2}\lambda_1 \, d\lambda_2\right] + \beta_{22}\int_0^{\lambda_2}\lambda_2 \, d\lambda_2$$

$$= \left[\beta_{11}\int_0^{\lambda_1}\lambda_1 \, d\lambda_1 + \beta_{12}\int_0^{\lambda_1\lambda_2} d(\lambda_1\lambda_2) + \beta_{22}\int_0^{\lambda_2}\lambda_2 \, d\lambda_2\right]$$

$$= \frac{1}{2}\beta_{11}\lambda^2 + \beta_{12}\lambda_1\lambda_2 + \frac{1}{2}\beta_{22}\lambda_2^2$$

6b) Derive the expression for torque in a singly excited system.

field energy & mechanical force



field produces a mechanical force F_f in the direction indicated which drives the mechanical system

the mechanical work done by the field when the armature moves a distance dx in positive direction is

$$dw_m = F_f \, dx$$

As per principle of energy conservation

mechanical energy output = electrical energy input — increase in field energy

$$F_f \, dx = i \, d\lambda - dw_f \quad —①$$

we know that $w_f' = i\lambda - w_f(\lambda, x)$

$$w_f = i\lambda - w_f'(i,x)$$

$$dw_f = d i\lambda - d w_f'(i,x)$$

$$dw_f = i \, d\lambda + \lambda \, di - \left[\frac{\partial w_f'}{\partial i} di + \frac{\partial w_f'}{\partial x} dx\right] \quad —②$$

..substituting eqn② in eqn①, we have

$$F_f \, dx = i \, d\lambda - \left[\lambda \, di + i \, d\lambda - \left(\frac{\partial w_f'}{\partial i} di + \frac{\partial w_f'}{\partial x} dx\right)\right]$$

$$F_f \, dx = -\lambda \, di + \frac{\partial w_f'}{\partial i} di - \frac{\partial w_f'}{\partial x} dx$$

$$F_f \, dx = \left[\frac{\partial w_f'}{\partial i} - \lambda\right] di + \frac{\partial w_f'}{\partial x} dx \quad —③$$

Because the incremental changes di and dx are independent and di is not present in the left hand side of eqn③ its coefficient on right hand side must be zero i.e

$$\frac{\partial w_f'}{\partial i} - \lambda = 0$$

$$\lambda = \frac{\partial w_f'}{\partial i}$$

$$\text{then } F_f = \frac{\partial w_f'(i,x)}{\partial x} \quad —④$$

The expression for mech. force developed applied when i is an independent if (λ, x) are taken as independent variables

$$w_f = w_f(\lambda, x)$$

$$dw_f = \frac{\partial w_f}{\partial \lambda} d\lambda + \frac{\partial w_f}{\partial x} dx \quad —⑥$$

substituting eqn⑤ in eqn①

$$F_f \, dx = i \, d\lambda - \frac{\partial w_f}{\partial \lambda} d\lambda - \frac{\partial w_f}{\partial x} dx$$

$$F_f \, dx = -\frac{\partial w_f}{\partial \lambda} \partial\lambda + \left(i - \frac{\partial w_f}{\partial \lambda}\right) d\lambda$$

Since dλ is independent differential, is not present on left hand side of

$$i - \frac{\partial W_f}{\partial \lambda} = 0$$

$$i = \frac{\partial W_f(\lambda, x)}{\partial \lambda}$$

then $F_f = -\frac{\partial W_f(\lambda, x)}{\partial x}$ —⑤

Determination of mechanical force

Non linear case

mechanical force is given by the partial derivatives of co-energy & energy

$$F_f = \left.\frac{\Delta W_f'}{\Delta x}\right|_{i = const} \quad \text{—⑥}$$

$$F_f = -\left.\frac{\Delta W_f}{\Delta x}\right|_{i = cost} \quad \text{—⑦}$$

these Two expressions will give slightly different Numerical values of F_f because of finite Δx. obviously F_f is the same in each case as $\Delta x \to 0$

Linear case $W_f'(i, x) = \int_0^i \lambda \, di = \int_0^1 L i \, di$

$$W_f'(i, x) = \frac{1}{2} L(x) i^2$$

$$F_f = \frac{\partial W_f'}{\partial x} = \frac{1}{2} i^2 \frac{\partial L(x)}{\partial x} \quad \text{—⑧}$$

from eqn⑧ it is obvious that the force acts in a direction

To increase the inductance of the exciting coil

$$W_f(\lambda, X) = \frac{1}{2} \frac{\lambda^2}{L(x)}$$

$$F_f = -\frac{\partial W_f}{\partial x} = \frac{1}{2}\left(\frac{\lambda}{L(x)}\right)^2 \frac{\partial L(x)}{\partial x}$$

$\because i = \lambda / L$

$$\Rightarrow W_f(\phi, x) = \frac{1}{2} R(x) \phi^2$$

$$F_f = -\frac{\partial W_f}{\partial x} = -\frac{1}{2} \phi^2 \frac{\partial R(x)}{\partial x}$$

$$\Rightarrow W_f(\lambda, x) = \frac{1}{2} \lambda i(x)$$

$$F_f = -\frac{\partial W_f}{\partial x} = -\frac{1}{2} \lambda \frac{\partial i(x)}{\partial x}$$

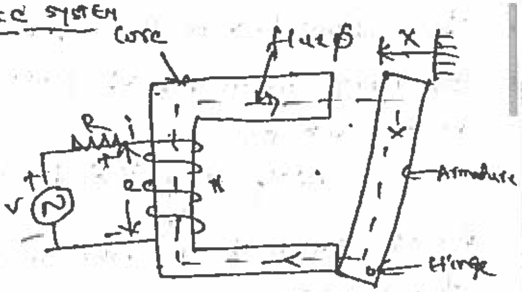
7a) for a singly excited system derive the expression for magnetic field energy stored?

ENERGY IN MAGNETIC SYSTEM

$\lambda = N\phi$

flux linkages

emf $e = \frac{d\lambda}{dt}$



$v = iR + e$

$= iR + \frac{d\lambda}{dt}$

The electric energy I/p into the Ideal coil due to the flow of current i in time dt is

$dW_e = e i \, dt$

Assuming for the time being that the armature is held fixed at position 'x', all the i/p energy stored in the magnetic field

$dW_e = dW_f = ei \, dt \; (ei\,dt)$

where dW_f is the change in field energy in time dt.

$dW_e = ei\,dt = \frac{d\lambda}{dt} i\,dt = i\,d\lambda$

$= i\,d(N\phi)$

$= Ni\,d\phi$

$= F\,d\phi$

$= dW_f$

$(F = Ni, \text{ mmf})$

Ans) •

The relation $i-\lambda$ or $F-\lambda$, non linear

The energy absorbed by the field for finite change in flux linkages

$$\Delta W_f = \int_{\lambda_1}^{\lambda_2} i(\lambda)\,d\lambda = \int_{\phi_1}^{\phi_2} F(\phi)\,d\phi$$

As the flux in the magnetic circuit undergoes a cycle $\phi_1 \to \phi_2 \to \phi_1$,

an irrecoverable loss in energy takes place due to hysteresis and eddy current in the iron,

The assumption renders the ideal coil and the magnetic circuit as a conservative system with energy interchange between themselves so that the net energy is conserved.

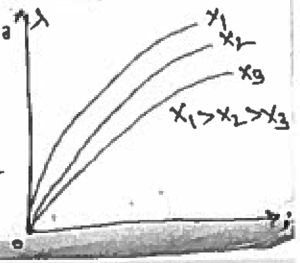
→ the energy absorbed by the magnetic system to establish flux ϕ (or flux linkages λ) from initial zero flux is

$$W_f = \int_0^{\lambda} i(\lambda)\,d\lambda = \int_0^{\phi} F(\phi)\,d\phi$$

⇒ The $i-\lambda$ relationship is magnetization curve which varies with the configuration variable x

→ The air gap b/w the armature and core varies with position x of the armature,

the total reluctance $\left(\frac{NI}{\phi}\right)$ of the magnetic path decreases as x increases



The i-λ relationship for various values of x is indicated in figure.

the relationship can be expressed as

$$i = i(\lambda, x)$$

if λ is the independent variable or as,

$$\lambda = \lambda(i, x)$$

if i is the independent variable

the field energy is in general, a function of two variables i.e. $W_f = W_f(\lambda, x)$ (or)

$$W_f = W_f(i, x)$$

→ A change in λ with fixed x (electro magnetic energy interchange

$$\left[\begin{array}{l} v = iR + \frac{d\lambda}{dt} \\ dw_e = i d\lambda = F d\phi = dW_f \end{array} \right]$$

→ if x is allowed to change with fixed λ, energy will interchange b/w the magnetic circuit and the mechanical system. (electro $\overset{+}{}$ mechanical) magnetic

→ As per $W_f = \int_0^\lambda i(\lambda)\, d\lambda = \int_0^\phi F(\phi)\, d\phi$

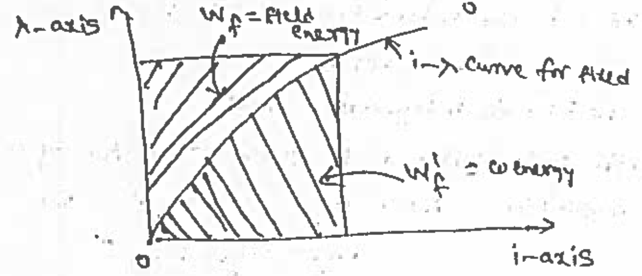
the field energy is the area b/w λ-axis and i-λ

A New term, co-energy is now defined as

$$W'_f(i, x) = i\lambda - W_f(\lambda, x)$$

wherein by expressing λ as λ(i, x), the independent variable of W'_f becomes i and x.

The co energy in fig is shown to be the complementary area of the i-λ curve. $W'_f = \int_0^i \lambda\, di$



Linear case

$$\omega_f = \frac{1}{2} i\lambda = \frac{1}{2} F\phi = \frac{1}{2} R\phi^2$$

∵ $R = F/\phi$ = reluctance of the magnetic circuit

∴ coil inductance $L = \frac{\lambda}{i}$

The field energy can be expressed as $\omega_f = \frac{1}{2}\frac{\lambda^2}{L}$

$$W_f(\lambda, x) = \frac{1}{2}\frac{\lambda^2}{L(x)}$$

The field energy is distributed throughout the space. Occupied by the field. Assuming no losses const permeability Energy density of the field $w_f = \int_0^B H\, dB = \frac{1}{2} HB = \frac{1}{2}\frac{B^2}{\mu}$

7b) write in brief about multiply excited magnetic system?

Ans) Multiply Excited Magnetic Field System – Singly-excited devices discussed earlier, are generally employed for motion through a limited distance or rotation through a prescribed angle.

Electro-mechanical transducers have the special requirement of producing an electrical signal proportional to forces or velocities or producing force proportional to electrical signal (current or voltage). Such transducers require two excitation one excitation establishes a magnetic field of specified strength while the other excitation produces the desired signal (electrical or mechanical).

Also continuous energy conversion devices motors and generators require multiple excitation.

Figure shows a magnetic field system with two electrical excitations one on stator and the other on rotor. The system can be described in either of the two sets of three independent variables ; (λ_1, λ_2, θ) or (i_1, i_2, θ). In terms of the first set

$$T_f = -\frac{\partial W_f(\lambda_1, \lambda_2, \theta)}{\partial \theta} \qquad (4.62)$$

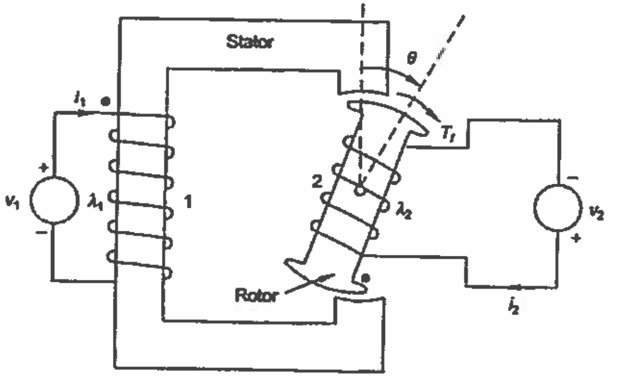


Fig. 4.15

where the field energy is given by

$$W_f(\lambda_1, \lambda_2, \theta) = \int_0^{\lambda_1} i_1\, d\lambda_1 + \int_0^{\lambda_2} i_2 d\lambda_2 \qquad (4.63)$$

Analogous to Eq. (4.28)

$$i_1 = \frac{\partial W_f(\lambda_1, \lambda_2, \theta)}{\partial \lambda_1}$$

$$i_2 = \frac{\partial W_f(\lambda_1, \lambda_2, \theta)}{\partial \lambda_2}$$

Assuming linearity

$$\lambda_1 = L_{11} i_1 + L_{12} i_2 \qquad (4.64a)$$

$$\lambda_2 = L_{21} i_1 + L_{22} i_2;\ (L_{12} = L_{21}) \qquad (4.64b)$$

Solving for i_1 and i_2 in terms of λ_1, λ_2 and substituting in Eq. (4.63) gives upon integration

$$W_f(\lambda_1, \lambda_2, \theta) = \frac{1}{2}\beta_{11}\lambda_1^2 + \beta_{12}\lambda_1\lambda_2 + \frac{1}{2}\beta_{22}\lambda_2^2 \qquad (4.65)$$

where

$$\beta_{11} = L_{22}/(L_{11}L_{22} - L_{12}^2)$$

$$\beta_{22} = L_{11}/(L_{11}L_{22} - L_{12}^2)$$

$$\beta_{12} = \beta_{21} = -L_{12}/(L_{11}L_{22} - L_{12}^2)$$

The Self- and mutual-inductance of the two exciting coils are functions are angle θ.

If currents are used to describe the system state

$$T_f = \frac{\partial W_f'(i_1, i_2, \theta)}{\partial \theta} \qquad (4.66)$$

where the co energy is given by

$$W_f'(i_1, i_2, \theta) = \int_0^{i_1} \lambda_1\, di_1 + \int_0^{i_2} \lambda_2\, di_2 \qquad (4.67)$$

In the linear case

$$W'_f(i_1, i_2, \theta) = \frac{1}{2} L_{11} i_1^2 + L_{12} i_1 i_2 + \frac{1}{2} L_{22} i_2^2$$

where inductances are functions of angle θ

8a) Explain the constructional details of dc generator?

Ans) Construction of DC machines : A D. C. machine consists of two main parts 1. Stationary part: It is designed mainly for producing a magnetic flux. 2. Rotating part: It is called the armature, where mechanical energy is converted into electrical (electrical generate) or conversely electrical energy into mechanical (electric into) Parts of a Dc Generator: 1) Yoke 2) Magnetic Poles a) Pole core b) Pole Shoe 3) Field Winding 4) Armature Core 5) Armature winding 6) Commutator 7) Brushes and Bearings The stationary parts and rotating parts are separated from each other by an air gap.

The stationary part of a D. C. machine consists of main poles, designed to create the magnetic flux, commutating poles interposed between the main poles and designed to ensure spark less operation of the brushes at the commutator and a frame / yoke. The armature is a cylindrical body rotating in the space between the poles and comprising a slotted armature core, a winding inserted in the armature core slots, a commutator and brush .

Yoke: 1. It saves the purpose of outermost cover of the dc machine so that the insulating materials get protected from harmful atmospheric elements like moisture, dust and various gases like SO2, acidic fumes etc. 2. It provides mechanical support to the poles. 3. It forms a part of the magnetic circuit. It provides a path of low reluctance for magnetic flux.

Poles: Each pole is divided into two parts a) pole core b) pole shoe Functions: 1. Pole core basically carries a field winding which is necessary to produce the flux. 2. It directs the flux produced through air gap to armature core to the next pole. 3. Pole shoe enlarges the area of armature core to come across the flux, which is necessary to produce larger induced emf. To achieve this, pole core has been given a particular shape.

Armature: It is further divided into two parts namely, (1) Armature core (2) Armature winding. Armature core is cylindrical in shape mounted on the shaft. It consists of slots on its periphery and the air ducts to permit the air flow through armature which serves cooling purpose. Functions: 1. Armature core provides house for armature winding i.e., armature conductors. 2. To provide a path of low reluctance path to the flux it is made up of magnetic material like cast iron or cast steel.

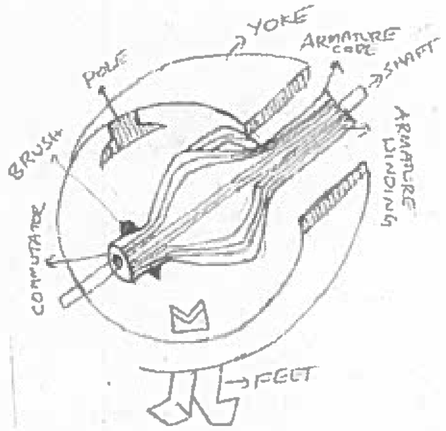
2. Armature winding: Armature winding is nothing but the inter connection of the armature conductors, placed in the slots provided on the armature core. When the armature is rotated, in case of generator magnetic flux gets cut by armature conductors and emf gets induced in them. Function: 1. Generation of emf takes place in the armature winding in case of generators. 2. To carry the current supplied in case of dc motors. 3. To do the useful work it the external circuit.

Field winding: The field winding is wound on the pole core with a definite direction. Functions: To carry current due to which pole core on which the winding is placed behaves as an electromagnet, producing necessary flux. As it helps in producing the magnetic field i.e. exciting the pole as electromagnet it is called 'Field winding' or 'Exciting winding'.

Commutator: The rectification in case of dc generator is done by device called as commutator. Functions: 1. To facilitate the collection of current from the armature conductors.

2. To convert internally developed alternating emf to in directional (dc) emf 3. To produce unidirectional torque in case of motor. Choice of material: As it collects current from armature, it is also made up of copper segments. It is cylindrical in shape and is made up of wedge shaped segments which are insulated from each other by thin layer of mica.

Brushes and brush gear: Brushes are stationary and rest on the surface of the Commutator. Brushes are rectangular in shape. They are housed in brush holders, which are usually of box type. The brushes are made to press on the commutator surface by means of a spring, whose tension can be adjusted with the help of lever. A flexible copper conductor called pigtail is used to connect the brush to the external circuit. Functions: To collect current from commutator and make it available to the external circuit.



8b) calculate the emf generated by a 4pole wave wound armature having 45 slots 18 conductors per slot .when driven at 1200 rpm the flux per pole is 0.016 wb.

Ans) E = (N P Ø Z)/(60 A)

Φ = flux in webers

n = armature speed in rpm

Z = total number of armature conductors

p = number of poles

a = number of parallel paths

for wave winding A =2

no of conductors =18 *45 = 810

emf generated = 0.016*810*1200*4\60*2

=62208/120 = 518.4 volts

9a) Explain magnetisation characteristics of a dc shunt generator ?

Ans : A DC generator, is an electrical machine that converts the mechanical energy of a prime mover (e.g. DC motor, AC induction motor or a turbine) into direct electrical energy. The generator shown in figure 1 is self-exciting. It uses the voltage Ea generated by the machine to establish the field current If, which in turn gives rise to the magnetic-field flux Φ. When the armature winding rotates in this magnetic field so to cut the flux, the voltage Ea is induced in the armature. This voltage is commonly referred to as the armature electromotive force or EMF. The induced EMF is proportional to the rate of cutting the flux and is given by emf formula.

E = (N P Ø Z)/(60 A)

where,

Φ = flux in webers

n = armature speed in rpm

Z = total number of armature conductors

p = number of poles

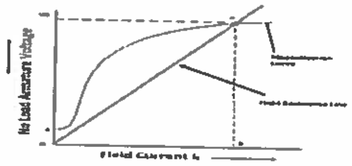
a = number of parallel paths

Critical Field Resistance:

It is that value of the field resistance at which the D.C. Shunt Generator will fail to excite.

It is that speed for which the given shunt field resistance becomes the critical field resistance. Critical field resistance is obtained by plotting the OCC as in fig.2 and determining the slope of the tangent to the linear position of the curve from the origin. While drawing the tangent, the initial position of the O.C.C is neglected.

Due to residual magnetism in the poles, a small amount of EMF is generated even when If = 0. Hence the curve starts a little way up. The slight curvature at the lower end is due to magnetic inertia. It has seen that the first part of the curve is practically straight. Hence the flux and consequently generated EMF is directly proportional to the exciting current. However at the higher flux densities where it is small iron path reluctance becomes appreciable and straight. Field windings are connected parallel to the armature, so it is called DC Shunt Generator. Due to residual magnetism some initial emf and hence some current will be generated. This current while passing into the field coils will strengthen the magnetism of poles. This will increase pole flux which will further increase the emf.the reinforcement of emf and flux proceeds till equilibrium is reached at some point in graph between IF and EG.



9B) Express emf equation of the dc generator?

Ans: EMF equation of a generator

Let P = number of poles

Ø = flux/pole in webers

Z = total number of armature conductors. = number of slots x number of conductors/slot

N = armature rotation in revolutions (speed for armature) per minute (rpm)

A = No. of parallel paths into which the 'z' no. of conductors are divided.

E = emf induced in any parallel path

Eg = emf generated in any one parallel path in the armature.

Average emf generated/conductor = dØ/dt volt

Flux current/conductor in one revolution dt = d x p In one revolution,

the conductor will cut total flux produced by all poles = d x p

No. of revolutions/second = N/60

Therefore, Time for one revolution, dt = 60/N second According to Faraday's laws of Electromagnetic Induction, emf dØ generated/conductor = dØ x p x N / 60 volts

This is emf induced in one conductor. For a simplex wave-wound generator No. of parallel paths = 2 No. of conductors in (series)in one path = Z/2

EMF generated/path = ØPN/60 x Z/2 = ØZPN/120 volt

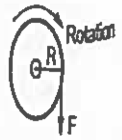
For a simple lap-wound generator Number of parallel paths = P Number of conductors in one path = Z/P

EMF generated/path = ØPN/60 (Z/P) = ØZN/60

A = 2 for simplex – wave winding A =P for simplex lap-winding .

10a) derive torque equation of a dc motor?

TORQUE EQUATION OF THE MOTOR



- The turning or twisting force about an axis called torque.
- Consider a wheel of radius "R" meters acted upon by a circumferential force F newtons
- The wheel is rotating at a speed of "N" rpm then its angular speed is,
- w(omega)= (2πN/60) rad/sec
- So workdone in one revolution is
- W= F * Distance travelled in one revolution = F * 2πR Joules
- P=Power developed = (workdone/time) =((F * 2πR)/time for 1 revolution)
- = ((F * 2πR)/(60/N)) = (F * R) * (2πN/60)
- P = T * w(omega) watts
- Where T = torque in Nm and w(omega)= angular speed in rad/sec
- Let Ta be the gross torque developed by the armature of the motor. It is also called armature torque
- The gross mechanical power developed in the armature is Eb*Ia. as seen from the power equation
- So if speed of the motor is N rpm then,
- Power in armature=armature torque * w(omega) i.e. EbIa= Ta * (2πN/60)
- But Eb in a motor is given by, Eb =(ΦZNP)/60A

$$\therefore \frac{\phi P N Z}{60\,A} \times I_a = T_a \times \frac{2\pi N}{60}$$

$$\therefore \boxed{T_a = \frac{1}{2\pi}\phi\, I_a \times \frac{PZ}{A} = 0.159\,\phi\, I_a \cdot \frac{PZ}{A} \quad \text{Nm}}$$

(10)b) A dynamo has a rated armature current at 250 A, what is the current per path of the armature if the armature winding is lap or wave connected ? the machine has 12 poles.

sol $I_a = 250\,A$, $P = 12$

$A_1 = 2$ number of parallel paths with wave winding

$A_2 = P$ number of parallel paths with lap winding

$I_1 = \frac{I_a}{A_1}$ Current per path in case of wave winding in Amps

$I_1 = \frac{250}{2} = 125\,A$

$I_2 = \frac{I_a}{A_2}$ Current per path in case of lap winding in Amps

$I_2 = \frac{250}{12} = 20.83$ Amps

11a) explain speed control of dc shunt motor?

Ans) The speed equation of a DC motor shows that,

$N \alpha Eb / \varphi$

or $N \alpha (V - IaRa)/ \varphi$

But the resistance of armature winding is small. Therefore the voltage drops IaRa will be negligible as compared to the external supply voltage V. Therefore, the expression for the speed can be approximated as follows:

$N \alpha V/\varphi$ (because V>> IaRa)

From this expression we can obtain the dc shunt motor speed control methods.

The speed is inversely proportional to flux φ.

It is directly proportional to armature voltage drop (IaRa).

It is directly proportional to applied Voltage V.

Speed Control of DC Shunt Motor:

So by varying one of these parameters, it is possible to change the speed of DC shunt motor. Depending on the parameter being controlled, methods of speed control of shunt DC Motor are classified as follows:

Flux Control Method:

speed control of dc shunt motor

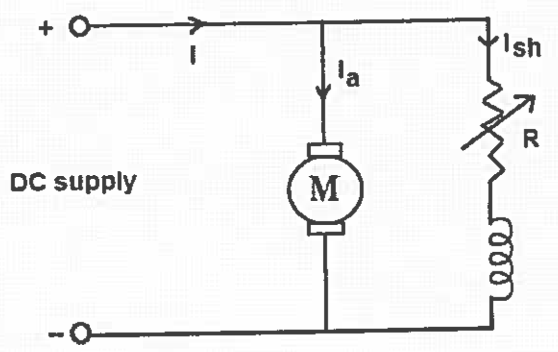
The set up for speed control of DC shunt motor using flux control technique is shown in the figure. In order to change the speed, we have to change flux. This can be achieved by changing the field current. The field current can be changed by changing the rheostat R connected in series with the field.At the time of starting the motor, we need to run the motor slowly, therefore, the flux should be maximum, because,

$N \alpha 1 / \varphi$

To obtain maximum flux at the start, the field current should be maximum at the time of starting. To obtain this, the value of rheostat (R) should be minimum.

The speed of DC shunt motor can be varied by varying the field current. As we increase the resistance R of the rheostat, the field current Ish decreases. So the flux φ decreases. This results in increasing the speed of the motor. As the R is increased, the speed increases.

We can use this technique to control motor speed above its rated value. It is the most commercial method. There is a limit to the maximum obtainable speed by this method due to poor commutation at week fluxes. Most common maximum to minimum speed ratio is 6:1.



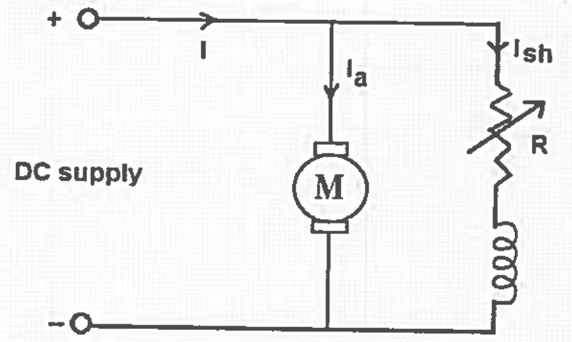
Speed Control of DC Shunt Motor by Armature:

speed control of dc shunt motor theory

The set up for speed control of DC shunt motor by armature voltage control method is shown in the figure. A rheostat is connected in series with armature winding. By varying the value of R we can vary the voltage across the armature.

Because speed N is directly proportional to armature voltage, it is possible to change the speed by changing the value of rheostat R.

We can use this technique to control motor speed below its rated value. But it is neither efficient nor economical method because, in this method, speed is reduced at the cost of power rheostat(Ia2R).



11b) explain in detail the losses in dc motor?

Ans:

- The various losses in a dc machine whether it is a motor or a generator are classified into three groups as:

1. Copper losses
2. Iron or core losses
3. Mechanical losses

Copper losses

- The copper losses are the losses taking place due to the current flowing in a winding.
- There are basically two windings in a dc machine namely armature winding and field winding.
- The copper losses are proportional to the square of the current passing through these windings.
- Armature copper loss = Ia'2 * Ra
- Shunt field copper loss = Ish'2 * Rsh

Series field copper loss = Ise'2 * Rse

Iron or core losses

- These losses are also called magnetic losses . these losses include hysteresis loss and eddy current loss
- The hysteresis loss is proportional to the frequency and the maximum flux density Bm in the air gap and is given by
- This loss is basically due to reversal of magnetisation of the armature core

$$\boxed{\text{Hysteresis loss} = \eta B_m^{1.6} f V \text{ watts}}$$

where

η = Steinmetz hysteresis coefficient

V = Volume of core in m^3

f = Frequency of magnetic reversals

eddy current loss

- The eddy current loss exists due to eddy currents.
- When armature core rotates, it cuts the magnetic flux and emf gets induced in the core.
- This induced emf sets up eddy currents which cause the power loss . this loss is given by
- These losses are almost constant

$$\boxed{\text{Eddy current loss} = K B_m^2 f^2 t^2 V \text{ watts}}$$

where

K = Constant

t = Thickness of each lamination

V = Volume of core

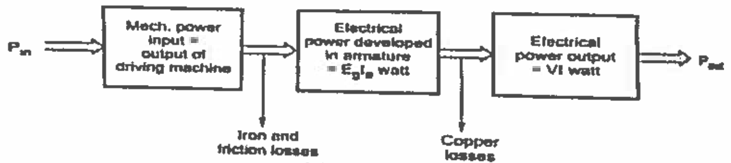
f = Frequency of magnetic reversals

TOTAL LOSSES

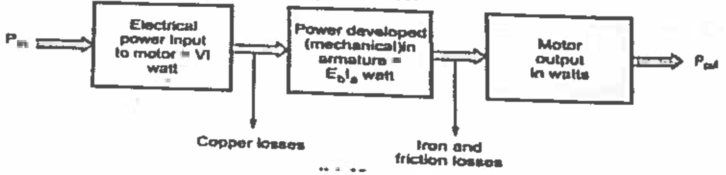
- For a dc machine

  Total losses = constant losses + variable losses

- The power flow and energy transformation diagrams at various stages,



(a) Generator



12a) derive the emf eqn of a transformer?

Ans: E.M.F. Equation of a Transformer Consider that an alternating voltage V1 of frequency f is applied to the primary . The sinusoidal flux ϕ produced by the primary can be represented as: ϕ=ϕm sinωt When the primary winding is excited by an alternating voltage V1, it is circulating alternating current, producing an alternating flux ϕ. ϕ - Flux ϕm - maximum value of flux

N1 - Number of primary turns

N2 - Number of secondary turns

f - Frequency of the supply voltage

E1 - R.M.S. value of the primary induced e.m.f

E2 - R.M.S. value of the secondary induced e.m.f

The instantaneous e.m.f. e1 induced in the primary is From Faraday's law of electromagnetic induction - The flux increases from zero value to maximum value ϕm in 1/4f of the time period that is in 1/4f seconds. The change of flux that takes place in 1/4f seconds = ϕm - 0 = ϕm weber.

$$\frac{d\phi}{dt} = \frac{dt}{1/4f} = 4f\phi_m \text{ wb/sec.}$$

Since flux ϕ varies sinusoidally, the R.m.s value of the induced e.m.f is obtained by multiplying the average value with the form factor

$$\text{Form factor of a sinwave} = \frac{\text{R.m.s value}}{\text{Average value}} = 1.11$$

R.M.S Value of e.m.f induced in one turns = $4\phi_m f \times 1.11$ Volts.

$= 4.44\phi_m f$ Volts.

R.M.S Value of e.m.f induced in primary winding = $4.44\phi_m f N1$ Volts.

R.M.S Value of e.m.f induced in secondary winding = $4.44\phi_m f N2$ Volts.

The expression of E_1 and E_2 are called e.m.f equation of a transformer

$V_1 = E_1 = 4.44\phi_m f N_1$ Volts.
$V_2 = E_2 = 4.44\phi_m f N_2$ Volts.

200turns

12b) 400/230V, 50 HZ

$E_1 = 400$ V

$E_2 = 230$ V

Turns on HV side

$N_1 = 200$ turns

Turns ratio $= \frac{N_1}{N_2}$

Transformation ratio

$$K = \frac{E_2}{E_1} = \frac{N_2}{N_1} = \frac{I_1}{I_2}$$

$$\frac{E_2}{E_1} = \frac{N_2}{N_1}$$

$$\frac{230}{400} = \frac{N_2}{200}$$

$$N_2 = \frac{230}{400} \times 200 = 115$$

$$\frac{N_2}{N_1} = \frac{115}{200} \text{ (Turns ratio)}$$

$$E_1 = 4.44 f \phi_m N_1$$

$$400 = 4.44 \times 50 \times \phi_m \times 200$$

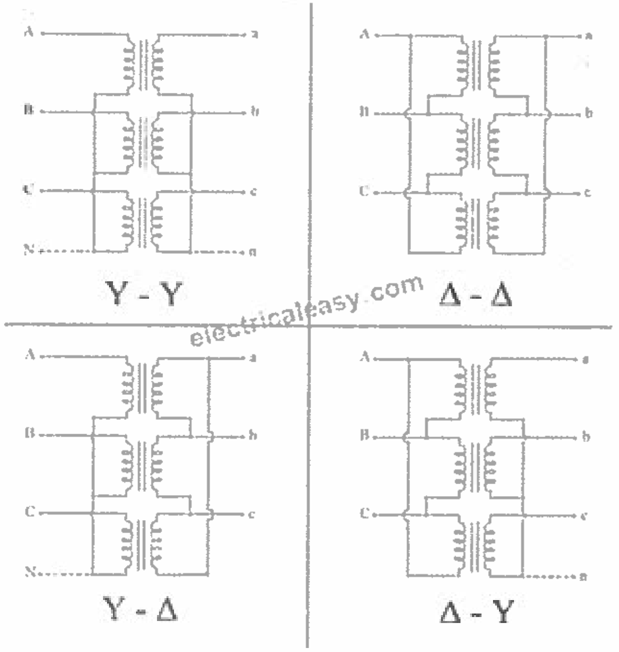
$$\phi_m = \frac{400}{4.44 \times 50 \times 200}$$

$$\phi_m = 0.0090 \text{ wb}$$

13a) draw the various types of three phase transformer connections and brief each one of them?

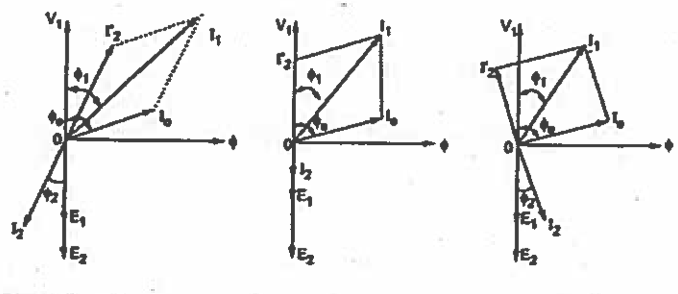
Ans: Three Phase Transformer Connections Three phase transformer connections In three phase system, the three phases can be connected in either star or delta configuration. In case you are not familiar with those configurations, study the following image which explains star and delta configuration. In any of these configurations, there will be a phase difference of 120° between any two phases.

Three Phase Transformer Connections Windings of a three phase transformer can be connected in various configurations as (i) star -star, (ii) delta-delta, (iii) star-delta, (iv) delta-star, (v) open delta and (vi) Scott connection. These configurations are explained below. Star-Star (Y-Y) • Star-star connection is generally used for small, high-voltage transformers. Because of star connection, number of required turns/phase is reduced (as phase voltage in star connection is 1/√3 times of line voltage only). Thus, the amount of insulation required is also reduced. • The ratio of line voltages on the primary side and the secondary side is equal to the transformation ratio of the transformers. • Line voltages on both sides are in phase with each other. • This connection can be used only if the connected load is balanced. Delta-Delta (Δ-Δ) • This connection is generally used for large, low-voltage transformers. Number of required phase/turns is relatively greater than that for star-star connection. • The ratio of line voltages on the primary and the secondary side is equal to the transformation ratio of the transformers. • This connection can be used even for unbalanced loading. • Another advantage of this type of connection is that even if one transformer is disabled, system can continue to operate in open delta connection but with reduced available capacity. Star-Delta OR Wye-Delta (Y-Δ) • The primary winding is star star (Y) connected with grounded neutral and the secondary winding is delta connected. • This connection is mainly used in step down transformer at the substation end of the transmission line. • The ratio of secondary to primary line voltage is 1/√3 times the transformation ratio. • There is 30° shift between the primary and secondary line voltages. Delta-Star OR Delta-Wye (Δ-Y) • The primary winding is connected in delta and the secondary winding is connected in star with neutral grounded. Thus it can be used to provide 3-phase 4-wire service. • This type of connection is mainly used in step-up transformer at the beginning of transmission line. • The ratio of secodary to primary line voltage is √3 times the transformation ratio. • There is 30° shift between the primary and secondary line voltages. Above transformer connection configurations are shown in the following figure. Open Delta (V-V) Connection Two transformers are used and primary and secondary connections are made as shown in t



13b) Draw and explain the phasor diagram of single phase transformer on load considering with winding resistance.

Ans) 1 Phasor Diagram i) Take (ϕ) flux as reference for all load iii) The load component I2', which is in anti-phase with I2 and phase of I2 is decided by the load. iv) Primary current I1 is vector sum of Io and I2' a) If load is Inductive, I2 lags E2 by ϕ2, shown in phasor diagram fig 2.7 (a). b) If load is resistive, I2 in phase with E2 shown in phasor diagram fig. 2.7 (b). c) If load is capacitive load, I2 leads E2 by ϕ2 shown in phasor diagram fig. 2.7 (c). For easy understanding at this stage here we assumed E2 is equal to V2 neglecting various drop.



(a) Inductive load (b) Resistive load (c) Capacitive load

Fig: 2.7.a

$I_1 \cong I_2'$

Balancing the ampere – turns

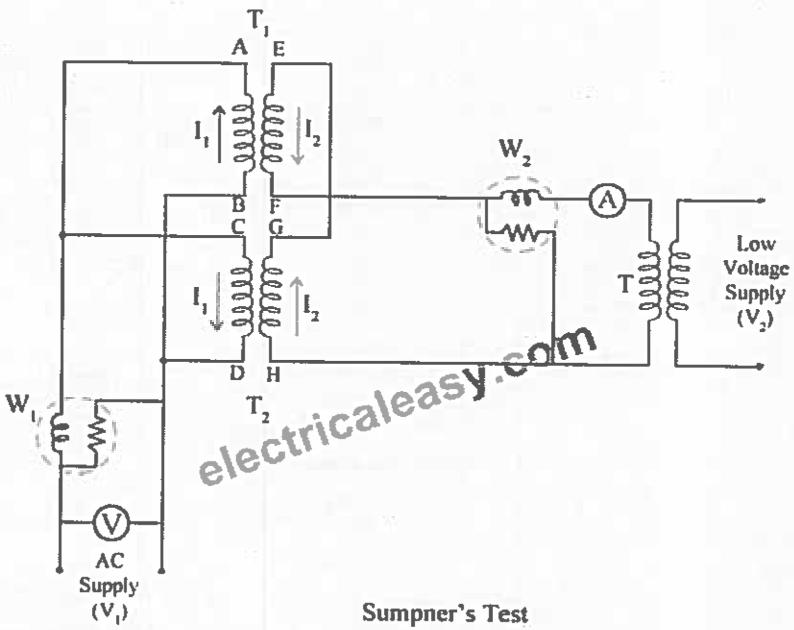
$N_1 I_2' = N_1 I_1 + N_2 I_2$

$\frac{I1}{I2} = \frac{N2}{N1} = K$

Effect of Winding Resistance In practical transformer it process its own winding resistance causes power loss and also the voltage drop. R1 – primary winding resistance in ohms. R2 – secondary winding resistance in ohms. The current flow in primary winding make voltage drop across it is denoted as I1R1 here supply voltage V1 has to supply this drop primary induced e.m.f E1 is the vector difference between V1 and I1R1. Similarly the induced e.m.f in secondary E2, The flow of current in secondary winding makes voltage drop across it and it is denoted as I2R2 here E2 has to supply this drop.

The vector difference between E2 and I2R2 (Assuming as purely resistive drop here.)

14a) draw the circuit diagram of sumpners test and derive the equation for efficiency of each transformer .



Sumpner's Test

Ans)

Sumpner's test or back to back test can be employed only when two identical transformers are available. Both transformers are connected to supply such that one transformer is loaded on another. Primaries of the two identical transformers are connected in parallel across a supply. Secondaries are connected in series such that emf's of them are opposite to each other. Another low voltage supply is connected in series with secondaries to get the readings, as shown in the circuit diagram .

In above diagram, T_1 and T_2 are identical transformers. Secondaries of them are connected in voltage opposition, i.e. E_{EF} and E_{GH}. Both the emf's cancel each other, as transformers are identical. In this case, as per superposition theorem, no current flows through secondary. And thus the no load test is simulated. The current drawn from V_1 is $2I_0$, where I_0 is equal to no load current of each transformer. Thus input power measured by wattmeter W_1 is equal to iron losses of both transformers.
i.e. iron loss per transformer Pi = $W_1/2$.

Now, a small voltage V_2 is injected into secondary with the help of a low voltage transformer. The voltage V_2 is adjusted so that, the rated current I_2 flows through the secondary. In this case, both primaries and secondaries carry rated current. Thus short circuit test is simulated and wattmeter W_2 shows total full load copper losses of both transformers.
i.e. copper loss per transformer P_{Cu} = $W_2/2$.

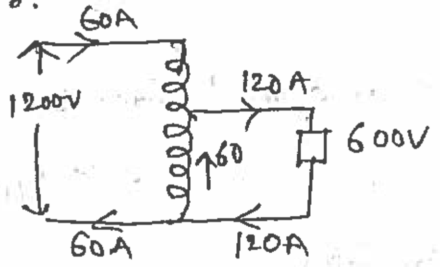
From above test results, the full load efficiency of each transformer can be given as -

$$\text{\% full load efficiency of each transformer} = \frac{\text{output}}{\text{output} + \frac{W_1}{2} + \frac{W_2}{2}} \text{ X } 100$$

14b) The primary and secodary voltages of an autotransformer are 1200v and 600v respectively .calculate the economy of cu when the secondary current is 120A .draw the circuit and show the current disturbtion in the winding .

④ b) The primary and secondary voltages of an auto transformer are 1200V and 600V respectively. calculate the economy of cu when the secondary current is 120 A. Draw the circuit and show the current distribution in the winding.

Sol



$$K = \frac{V_2}{V_1} = \frac{\cancel{600}}{\cancel{1200}_{2}} = \frac{1}{2} = 0.5$$

$$I_1 = K I_2 = 0.5 \times 120A = 60A$$

$$\text{saving} = K w_0 = 0.5 \times W_0 = 0.5 W_0$$

$\therefore$ per centage of saving $= 0.5 \times 100$

$= 50\%$.

Ans)

15a) A 10 kva 500/250 v 50 hz single phase transformer gave the following test data

Oc test (LV test):250 V,1.0A,80W

SC TEST (H V SIDE):25 V,12A,100W where LV refers to the low voltage and HV side refers to high voltage side .determine the follwing:

Ans)

Transformer rating $= 10 \text{KVA}$

$E_1 = 500\text{V}, \quad E_2 = 250\text{V}$

oc Test (Lv side) : $\underset{V_2}{250\text{V}}$ $\underset{I_0}{1.0\text{A}}$ $\underset{W_0}{80\text{W}}$

sc Test (Hv side) : $\underset{V_{(Sc)}}{25\text{V}}$, $\underset{I_{(sc)}}{12\text{A}}$, $\underset{W_c}{100\text{W}}$

open circuit test performance on LV side

$I_0 = 1.0\text{A}, \quad I_W = \frac{W_0}{V_2} = \frac{80}{250}$

$= 0.32\text{A}$

$$I_{mag} = \sqrt{I_0^2 - I_W^2}$$

$$= \sqrt{(1.0)^2 - (0.32)^2}$$

$$= \sqrt{1 - 0.1024}$$

$$= \sqrt{0.8976} = 0.9474\text{A}$$

$$R_0 = \frac{V_2}{I_W} = \frac{250}{0.32} = 781.25\,\Omega$$

$$X_0 = \frac{V_2}{I_{mag}} = \frac{250}{0.9474} = 263.8800\,\Omega$$

short circuit test performed on HV side

$Z_{ep} = \frac{V_1(sc)}{I_1(sc)} = \frac{25}{12} = 2.0833 \Omega$

$R_{ep} = \frac{W_c}{(I_1(sc))^2} = \frac{100}{(12)^2} = 0.6944 \Omega$

$X_{ep} = \sqrt{(Z_{ep})^2 - (R_{ep})^2} = \sqrt{(2.0833)^2 - (0.6944)^2}$

$= \sqrt{4.3401 - 0.4821}$

$= \sqrt{3.858}$

$= 1.9641$

Transformation ratio $= k = \frac{E_2}{E_1} = \frac{250}{500}$

$= \frac{1}{2} = 0.5$

when pf $\cos\phi = 0.8$ lag

$\sin\phi = 0.6$

$V_2 = E_2 - I_2 R_{es} \cos\phi - I_2 X_{es} \sin\phi$

full load current on HV side

$= \frac{10 \text{ kVA}}{500} = \frac{10 \times 1000}{500}$

$= 20 A$

parameters referred to LV side

$R_{es} = k^2 R_{ep} = (0.5)^2 \, 0.6944 = 0.1736$

$X_{es} = k^2 X_{ep} = (0.5)^2 \, 1.9641 = 0.4910$

$V_2 = E_2 - I_2 R_{es} \cos\phi - I_2 X_{es} \sin\phi$

$= 250 - \overset{20}{40} \times 0.1736 \times 0.8 - 40 \times 0.4910 \times 0.6$

full load current on LV side $= I_2 = \frac{10000}{250}$

$= 40$

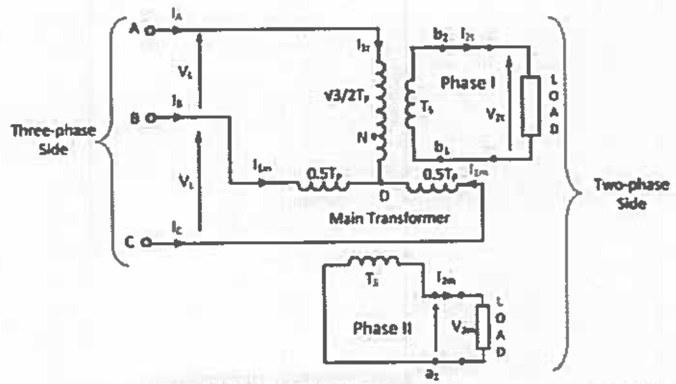
$V_2 = 232.66 V$

15 b) explain how the scott connection can be used to ontain two phase supply from a three phaswe supply?

Ans)

The Scott-T Connection is the method of connecting two single phase transformer to perform the 3-phase to 2-phase conversion and vice-versa. The two transformers are connected electrically but not magnetically.One of the transformers is called the main transformer, and the other is called the auxiliary or teaser transformer.

The figure below shows the Scott-T transformer connection. The main transformer is centre tapped at D and is connected to the line B and C of the 3-phase side. It has primary BC and secondary a_1a_2. The teaser transformer is connected to the line terminal A and the centre tapping D. It has primary AD and the secondary b_1b_2



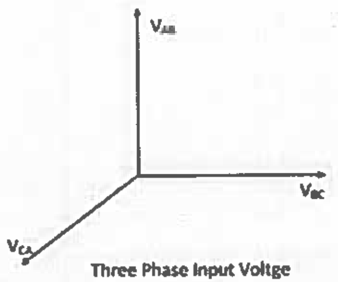
Scott connection Transformers

The identical, interchangeable transformers are used for Scott-T connection in which each transformer has a primary winding of T_p turns and is provided with tapping at $0.289T_P$, $0.5T_P$and 0.866 T_P.

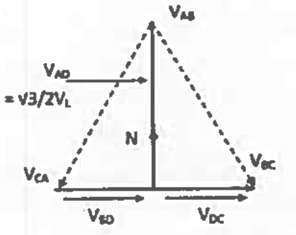
Phasor Diagram of Scott Connection Transformer

The line voltages of the 3-phase system V_{AB}, V_{BC}, and V_{CA} which are balanced are shown in the figure below. The same voltage is shown as a closed equilateral triangle.The figure below shows the primary windings of the main and the teaser transformer.

$|V_{AB}| = |V_{BC}| = |V_{CA}|=|V_L|$



Three Phase Input Voltge



Voltage on transformer Primary Winding

Circuit Globe

The D divides the primary BC of the main transformers into two halves and hence the number of turns in portion BD = the number of turns in portion DC = $T_p/2$.The voltage V_{BD} and V_{DC} are equal, and they are in phase with V_{BC}.

$$V_{BD} = V_{DC} = \frac{1}{2}V_{BC} = \frac{1}{2}V_L < 0°$$

The voltage between A and D is

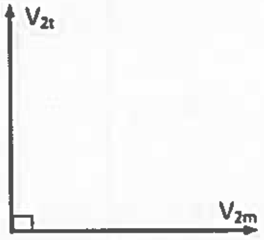
$$V_{BD} = V_{DC} = \frac{1}{2}V_{BC} = \frac{1}{2}V_L < 0°$$

$$V_{AD} = V_{AB} + V_{BD}$$

$$V_{AD} = V_L\left(-\frac{1}{2} + j\frac{\sqrt{3}}{2}\right) + \frac{1}{2}V_L$$

$$V_{AD} = V_L\left(j\frac{\sqrt{3}}{2}\right) = 0.866V_L < 90°$$

The teaser transformer has the primary voltage rating that is √3/2 or 0.866 of the voltage ratings of the main transformer. Voltage V_{AD} is applied to the primary of the teaser transformer and therefore the secondary of the voltage V_{2t} of the teaser transformer will lead the secondary terminal voltage V_{2m} of the main transformer by 90° as shown in the figure below.



Circuit Globe

$$\frac{V_{S_1}}{V_{AD}} = \frac{T_S}{T_{AD}}$$

$$V_{2t} = \frac{T_S}{T_{AD}}V_{AD} = \frac{T_S}{\frac{\sqrt{3}}{2}T_P} \times \frac{\sqrt{3}V_L}{2}$$

Then,

$$\frac{T_S}{T_P}V_L = v_{2m}$$

For keeping the voltage per turn same in the primary of the main transformer and the primary of the teaser transformer, the number of turns in the primary of the teaser transformer should be equal to $\sqrt{3}/2T_P$.

Thus, the secondaries of both transformers should have equal voltage ratings.The V_{2t} and V_{2m} are equal in magnitude and 90° apart in time; they result in the balanced 2-phase system.

Position of Neutral Point N

The primary of the two transformers may have a four wire connection to a 3-phase supply if the tapping N is provided on the primary of the teaser transformer such that

The voltage across AN = V_{AN} = phase voltage = $V_L/\sqrt{3}$.

Since the voltage across the portion AD.

$$V_{AD} = \frac{\sqrt{3}}{2}V_L$$

the voltage across the portion ND

$$V_{ND} = V_{AD} - V_{AN} = \frac{\sqrt{3}}{2}V_L - \frac{V_L}{\sqrt{3}} = \frac{V_L}{2\sqrt{3}}$$

The same voltage turn in portion AN, ND and AD are shown by the equations,

$$T_{AN} = \frac{T_P}{\sqrt{3}} = 0.577T_p$$

$$T_{ND} = \frac{T_P}{2\sqrt{3}} = 0.288T_p$$

$$T_{ND} = \frac{\sqrt{3}T_P}{2} = 0.866T_p$$

$$\frac{T_{AN}}{T_{ND}} = \frac{T_P}{\sqrt{3}} + \left(\frac{T_P}{2\sqrt{3}}\right) = 2$$

The equation above shows that the neutral point N divides the primary of the teaser transformer in ratio.

AN : ND = 2 : 1

Course Instructor

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19/5/22

HOD-EEE

(Dr. R.S.R. Krishnam naidu)

Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2021 - 2022
Course Code	20EC304	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Random Variables and Stochastic Process**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	State Bays theorem.	20EC304.1	L1
2	What does the Chebychev's inequality states?	20EC304.2	L1
3	List four properties of Joint distribution function.	20EC304.3	L1
4	Differentiate between Autocorrelation and Cross-correlation function.	20EC304.4	L1
5	Define narrow Pass Process.	20EC304.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Let XX be a random variable with PDF given by $f_X(x) = \begin{cases} cx^2 & \lvert x \rvert \le 1 \\ 0 & \text{otherwise} \end{cases}$ Find constant c, Variance (X) and P(X ≥ ½)	6M	20EC304.1	L2
6 (b)	Elaborate on Discrete and continuous sample spaces.	6M	20EC304.1	L2
	OR			
7 (a)	Let X be a continuous random variable with PDF $f_X(x) = \begin{cases} 4x^3 & 0 < x \le 1 \\ 0 & \text{otherwise} \end{cases}$ Find P(X ≤ 2/3 | X >1/3)	6M	20EC304.1	L2
7 (b)	Elaborate on Conditional density and their properties.	6M	20EC304.1	L2
8 (a)	State and prove Chebychev's inequality	6M	20EC304.2	L3
8 (b)	If X be a random variable and P(X-x) is the probability mass function given as below table	6M	20EC304.2	L3

X	0	1	2	3	4	5	6	7
P(X-x)	0	k	2k	3k	4k	K^2	$2K^2$	$7K^2 + k$

Find the value of k and P(X≤6)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
	OR			
9 (a)	Let Y=2X+3, If the random variable is uniformly distributed over [-1, 2], determine fy(y)	6M	20EC304.2	L3
9 (b)	Explain about the characteristic function and state its properties	6M	20EC304.2	L2
10	State and explain the properties of Marginal distribution functions, Conditional distribution and density function.	12M	20EC304.3	L3
	OR			
11 (a)	Consider two random variables X and Y with joint Probability Mass Function given in Table below.	6M	20EC304.3	L3

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	$Y = 2$	$Y = 4$	$Y = 5$
$X = 1$	$\frac{1}{12}$	$\frac{1}{24}$	$\frac{1}{24}$
$X = 2$	$\frac{1}{6}$	$\frac{1}{12}$	$\frac{1}{8}$
$X = 3$	$\frac{1}{4}$	$\frac{1}{8}$	$\frac{1}{12}$

Find P(X ≤ 2, Y ≤ 4) , P(Y =2| X =1) and check X and Y are independent.

11 (b)	State and prove central limit theorem for equal distributions case.	6M	20EC304.3	L3
12 (a)	Explain about Second-order and wide-sense stationary process.	6M	20EC304.4	L2
12 (b)	Derive the Relationship between cross-power density spectrum and cross-correlation function.	6M	20EC304.4	L4
	OR			
13 (a)	Explain how random processes are classified with neat sketches.	6M	20EC304.4	L2
13 (b)	Derive the relationship between power spectral density and autocorrelation function.	6M	20EC304.4	L4
14 (a)	A random processes X(t)= Asin(ωt+θ) , where A , ω are constants and θ is a uniformly distributed random variable on the interval (-π ,π) .find average power?	6M	20EC305.5	L2
14 (b)	Derive the relation between input PSD and output PSD of an LTI system.	6M	20EC304.5	L4
	OR			
15 (a)	Explain the following i) Noise Figure ii) Noise Sources.	6M	20EC304.5	L2
15 (b)	Derive the expression for average cross power between two random process X(t) and Y(t).	6M	20EC304.5	L4

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ANSWE KEY AND SCHEME OF EVALUATION:

PART-A:

1. State Baye's theorem - 2m

2. What does the chebychev's inequality statement - 2m

3. List four properties of joint distribution function - 2m
each property - 1/2 mark.

4. Differentiate Autocorrelation function - 1 m
Cross Correlation - 1 m

5. Define Narrow Band process - 2m.

PART-B:

6. (a) find constant c - 3m } (6m)
Variance (X) - 3m

(b) Discrete sample space - 3m } (6m)
Continuous Sample Space - 3m

7. (b) Properties of Conditional density function - 6m

(a) Problem - 6m.

(8) (a) State & prove Chebychev's inequality - 6m

(b) Problem, find k - 3m
find $P(X \leq 6)$ - 3m } - 6m.

(9) a) Problem - 6m.

b) characteristic function → 2m
Properties → 4m } 6m.

(10) marginal distribution function → 4m
density function → 4m
conditional distribution properties → 4m } 12m.

(11) (a) Problem, find $P(X \leq 2, Y \leq 4)$ → 3m
$P(Y=2/X=1)$ → 3m } 6m.

(b) Central limit theorem - 6m

(12) (a) Second order wide sense stationary - 6m

(b) Derivation cpds & ccf → 6m

(13) (a) Classification of Random process → 6m
each R.P → $1\frac{1}{2}$ marks
$4 \times 1\frac{1}{2} = 6$m

(b) Derivation of PSD & ACF → 6m

(14) (a) Problem - 6m.

(b) Derivation of input PSD & output PSD → 6m.

(15) (a) Noise figure → 3m
Noise Sources → 3m } 6m.

(b) Derivation of average cross power b/w two r.p $x(t)$ & $y(t)$. → 6m.

PART – A

1. State Baye's Theorem.

Ans: **Baye's Theorem:**

If $B_1, B_2, B_3, \cdots B_n$ are n mutually exclusive and exhaustive events such that $P(B_i) > 0$ ($i = 1, 2, 3, \cdots n$) in a sample space S and A is any other event in S intersecting every B_i (i.e A can only occur in combination with any one of the events $B_1, B_2, B_3, \cdots B_n$) such that $P(A) > 0$ then the conditional probability of B_i given A is

$$P\left(\frac{B_i}{A}\right) = \frac{P(B_i)\, P(A/B_i)}{\sum_{i=1}^{n} P(B_i)\, P(A/B_i)}$$

2. What does the Chebychev's inequality states?

Ans: **Chebychev's Inequality :-**

If x is a random variable with mean and variance σ^2 then (i) $P\{|x-\mu| \geq k\sigma\} \leq \frac{1}{k^2}$ and

(ii) $P\{|x-\mu| < k\sigma\} \geq 1 - \frac{1}{k^2}$

③ List four properties of joint distribution function

Ans: The properties of joint distribution function for two random variables X and Y are given as

1. $F_{X,Y}(-\infty,-\infty)=0$

   $F_{X,Y}(x,-\infty)=0$

   $F_{X,Y}(-\infty,y)=0$

2. $F_{X,Y}(\infty,\infty)=1$

3. $0 \le F_{XY}(x,y) \le 1$

4. $F_{XY}(x,y)$ is monotonic and non-decreasing function of both x and y.

5. The probability of the joint event $\{x_1 < X \le x_2,\ y_1 < Y \le y_2\}$ is given by $P\{x_1 < X \le x_2,\ y_1 < Y \le y_2\} = F_{XY}(x_2,y_2) + F_{XY}(x_1,y_1) - F_{XY}(x_1,y_2) - F_{XY}(x_2,y_1)$

6. The marginal distribution functions are given by

$$F_{X,Y}(x,\infty) = F_X(x) \text{ and } F_{X,Y}(\infty,y) = F_Y(y)$$

④ Differentiate between Autocorrelation and Cross-correlation

Ans: Auto Correlation:

Consider a random process $x(t)$. Let X_1 and X_2 be two random variables defined at times t_1 and t_2 with density function $f_X(x_1,x_2;t_1,t_2)$. The correlation of X_1 and X_2, $E[X_1X_2] = E[x(t_1)\,x(t_2)]$ is auto correlation.

$$R_{XX}(t_1,t_2) = \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} x_1 x_2 f_X(x_1,x_2;t_1,t_2)\,dx_1\,dx_2$$

Cross Correlation:

Consider two random processes $X(t)$ and $Y(t)$ defined with random variables X and Y at times t_1 and t_2 respectively. Then the correlation of X and Y

$E[XY] = E[X(t_1)\,Y(t_2)]$ is called the cross correlation function.

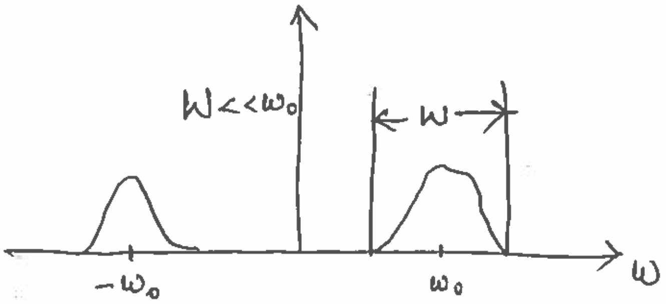
$$R_{XY}(t_1,t_2) = \int_{-\infty}^{\infty} xy f_{XY}(x,y;t_1,t_2)\,dxdy$$

⑤ Define Narrow Band Process

Ans:

Narrow Band Process:

When the bandwidth 'W' of band limited process is very less than 'ω_0' i.e if $W << \omega_0$, where ω_0 is the frequency at which power spectrum is maximum, then it is said to be narrow band process.



PART- B

6) a) Let X be a random Variable with PDF given by

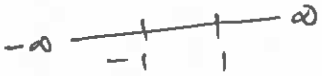
$$f_X(x) = \begin{cases} cx^2, & |x| \leq 1 \\ 0, & \text{otherwise} \end{cases}$$ Find Constant 'c'. variance 'x' and $P(X \geq \frac{1}{2})$

Ans: Given $f_X(x) = \begin{cases} cx^2, & |x| \leq 1 \\ 0, & \text{otherwise} \end{cases}$

i.e $|x| \leq 1 \Rightarrow -1 \leq x \leq 1$

To find Constant c

we know that $\int_{-\infty}^{\infty} f_X(x)\,dx = 1$



$$\int_{-\infty}^{-1} f_X(x)\,dx + \int_{-1}^{1} f_X(x)\,dx + \int_{1}^{\infty} f_X(x)\,dx = 1$$

$$0 + \int_{-1}^{1} cx^2\,dx + 0 = 1$$

$$c\int_{-1}^{1} x^2\,dx = 1$$

$$c\left[\frac{x^3}{3}\right]_{-1}^{1} = 1$$

$$\frac{c}{3}\left[1-(-1)^3\right] = 1$$

$$\frac{c}{3}[2] = 1 \Rightarrow \boxed{c = \frac{3}{2}} \text{ (or) } 1.5$$

Now find Mean $E[x] = \int_{-\infty}^{\infty} x\, f_X(x)\,dx$

$$= \int_{-1}^{1} x \cdot cx^2\,dx$$

$$= c\int_{-1}^{1} x^3\,dx$$

$$= c\left[\frac{x^4}{4}\right]_{-1}^{1}$$

$$= \frac{c}{4}\left[1-(-1)^4\right]$$

$$E[x] = \frac{c}{4}(1-1] = 0$$

$$\therefore \boxed{E[x] = 0}$$

Now find Variance 'x', $\sigma^2 = E[x^2] - [E[x]]^2 \rightarrow ①$

Now find $E[x^2] = \int_{-\infty}^{\infty} x^2 f_x(x)\, dx$

$$= \int_{-1}^{1} x^2 \cdot c x^2\, dx$$

$$= c \int_{-1}^{1} x^4\, dx$$

$$= \frac{3}{2} \left[\frac{x^5}{5}\right]_{-1}^{1}$$

$$= \frac{3}{10}\left[1-(-1)^5\right] = \frac{3}{\cancel{10}_5}[\cancel{2}]$$

$$\therefore E[x^2] = \frac{3}{5}$$

from ①

$\therefore$ Variance 'x' $= \frac{3}{5} - 0^2$

i.e $\boxed{\sigma^2 = \frac{3}{5}}$ //

Now find

$P(X \geq \frac{1}{2}) = P(X \geq 0.5) = \int_{0.5}^{1} c x^2\, dx$

$$= c \int_{0.5}^{1} x^2\, dx$$

$$= \frac{\cancel{3}}{2}\left[\frac{x^3}{\cancel{3}}\right]_{0.5}^{1}$$

$$= \frac{1}{2}\left[1-(0.5)^3\right]$$

$$= \frac{1}{2}\left[1-0.125\right]$$

$$\boxed{P(X \geq \tfrac{1}{2}) = 0.4375}$$ //

6) b) Elaborate on Discrete and Continuous Sample Spaces.

Ans :-

Discrete Sample Space :-

A Sample space is said to be discrete and finite of the set

Eg: $S = \{1, 2, 3, 4\}$ is a finite set

* If the set in the sample Space have infinite no-of elements then the sample space is discrete and finite

Eg: $S = \{1, 2, 3, \dots\}$ is infinite set.

Continuous Sample Space :-

If the sample space contains infinite number of values with continuous values with a given range then it is called continuous sample space

Eg: $S = \{0, 1, 2, 3, \dots 100\}$. The set of numbers 0 to 100 the elements are infinite and continuous.

7 (a) Let x be a continuous random Variable with PDF

$$f_x(x) = \begin{cases} 4x^3 & , 0<x<1 \\ 0 & , \text{otherwise} \end{cases} \quad \text{Find } P\left(x \le \frac{2}{3} / x > \frac{1}{3}\right).$$

Ans: Given $f_x(x) = \begin{cases} 4x^3, & 0<x<1 \\ 0, & \text{otherwise} \end{cases}$

$$P\left(X \le \frac{2}{2} / x > \frac{1}{3}\right) = \frac{P\left(x \le \frac{2}{3}, x > \frac{1}{3}\right)}{P\left(x > \frac{1}{3}\right)}.$$

$$P\left(x \le \frac{2}{3}, x > \frac{1}{3}\right) = \int_{1/3}^{2/3} 4x^3 \, dx$$

$$= 4 \int_{1/3}^{2/3} x^3 \, dx$$

$$= \cancel{4} \left[\frac{x^4}{\cancel{4}}\right]_{1/3}^{2/3}$$

$$= \left(\frac{2}{3}\right)^4 - \left(\frac{1}{3}\right)^4$$

$$= 0.189 - 0.0119 = 0.1771$$

$$P\left(x > \frac{1}{3}\right) = \int_{1/3}^{1} 4x^3 \, dx = \cancel{4}\left[\frac{x^4}{\cancel{4}}\right]_{1/3}^{1}$$

$$= \left[1 - \left(\frac{1}{3}\right)^4\right]$$

$$= \left[1 - \cancel{0}.0119\right]$$

$$P\left(x > \frac{1}{3}\right) = 0.9881$$

$$\therefore P(X \leq \tfrac{2}{3} / X > \tfrac{1}{3}) = \frac{0.1771}{0.9881}$$

$$\boxed{P(X \leq \tfrac{2}{3} / X > \tfrac{1}{3}) = \underline{0.179}}$$

7(b) Elaborate on Conditional density and their properties.

Ans:

Conditional density function:

Conditional density function of a random variable X as the derivative of the conditional distribution function If we denote this density by $f_X(x/B)$, then

$$f_X(x/B) = \frac{d}{dx}\left[F_X(x/B)\right]$$

If $F_X(x/B)$ contains step discontinuities, as when X is a discrete or mixed random variable, we assume that impulse functions are present in $f_X(x/B)$ to account for the derivatives at the discontinuities.

Properties of Conditional density:-

1) $f_X(x/B) \geq 0$

2) $\int_{-\infty}^{\infty} f_X(x/B)\,dx = 1$

3) $F_X(x/B) = \int_{-\infty}^{x} f_X(x/B)\,dx$

4) $P(x_1 \leq X \leq x_2/B) = \int_{x_1}^{x_2} f_X(x/B)\,dx$.

8 a) State and Prove Chebychev's Inequality

Ans : <u>Statement</u> : If K is a random variable with mean μ and variance σ^2 then (i) $P\{|x-\mu| \geq k\sigma\} \leq \frac{1}{k^2}$ and

(ii) $P\{|x-\mu| < k\sigma\} \geq 1-\frac{1}{k^2}$

Proof :- Given that

X is a random variable with mean μ and variance σ^2

we know that

$$\sigma^2 = V(x) = E\left[(x-\mu)^2\right]$$

$$= \int_{-\infty}^{\infty} (x-\mu)^2 f_X(x)\,dx$$

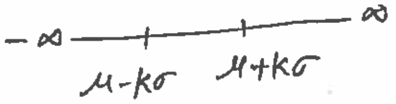
$\because |x-\mu| = k\sigma$

$x-\mu = \pm k\sigma$

$x = \mu \pm k\sigma$

i.e $x = \mu + k\sigma$

$x = \mu - k\sigma$



$$\sigma^2 = \int_{-\infty}^{\mu-k\sigma} (x-\mu)^2 f_X(x)\,dx + \int_{\mu-k\sigma}^{\mu+k\sigma} (x-\mu)^2 f_X(x)\,dx + \int_{\mu+k\sigma}^{\infty} (x-\mu)^2 f_X(x)\,dx$$

Since $\int_{\mu-k\sigma}^{\mu+k\sigma} (x-\mu)^2 f_X(x) \geq 0$ (always +ve)

Then $$\sigma^2 \geq \int_{-\infty}^{\mu-k\sigma} (x-\mu)^2 f_X(x)\,dx + \int_{\mu+k\sigma}^{\infty} (x-\mu)^2 f_X(x)\,dx \rightarrow ①$$

for the first integral in eqn ①

$x \leq \mu - K\sigma$

$x - \mu \leq -K\sigma$

multiply on b.s by '−'

$-(x-\mu) \geq K\sigma$

Squaring on both sides

$[-(x-\mu)^2] \geq (K\sigma)^2$

$(x-\mu)^2 \geq K^2\sigma^2 \rightarrow ②$

for the second integral in eq ①

$x \geq \mu + K\sigma$

$x - \mu \geq K\sigma$

squaring on b.s

$(x-\mu)^2 \geq K^2\sigma^2 \rightarrow ③$

Using equations ② & ③ in eqn ①, we get

$$\sigma^2 \geq \int_{-\infty}^{\mu-K\sigma} K^2\sigma^2 f_X(x)dx + \int_{\mu+K\sigma}^{\infty} K^2\sigma^2 f_X(x)dx$$

$$\sigma^2 \geq K^2\sigma^2 \left[\int_{-\infty}^{\mu-K\sigma} f_X(x)dx + \int_{\mu+K\sigma}^{\infty} f_X(x)dx\right]$$

$$1 \geq K^2 \left[\int_{-\infty}^{\mu-K\sigma} f_X(x)dx + \int_{\mu+K\sigma}^{\infty} f_X(x)dx\right]$$

$$\frac{1}{K^2} \geq \int_{-\infty}^{\mu-K\sigma} f_X(x)dx + \int_{\mu+K\sigma}^{\infty} f_X(x)dx$$

$$\frac{1}{K^2} \geq P[x \leq \mu - K\sigma] + P[x \geq \mu + K\sigma]$$

$$\frac{1}{K^2} \geq P[x-\mu \leq -K\sigma] + P[x-\mu \geq K\sigma]$$

$$\frac{1}{K^2} \geq P[-(x-\mu) \geq K\sigma] + P[x-\mu \geq K\sigma]$$

$$\frac{1}{k^2} \geq P[|x-\mu| \geq k\sigma]$$

i.e $P[|x-\mu| \geq k\sigma] \leq \frac{1}{k^2} \rightarrow$ ④

Now multiply '−' on both sides eqn ④

$$-P[|x-\mu| \geq k\sigma] \geq -\frac{1}{k^2}$$

adding 1 on both sides

$$1-P[|x-\mu| \geq k\sigma] \geq 1-\frac{1}{k^2}$$

i.e $P[|x-\mu| < k\sigma] \geq 1-\frac{1}{k^2}$

Hence proved.

8) b) If X be a random variable and $P(X=x)$ is the probability mass function given as below table

X	0	1	2	3	4	5	6	7
$P(X=x)$	0	k	2k	3k	4k	k^2	$2k^2$	$7k^2+k$

Find the value of k and $P(X \leq 6)$

Ans:- Given table

X	0	1	2	3	4	5	6	7
$P(X=x)$	0	k	2k	3k	4k	k^2	$2k^2$	$7k^2+k$

we know that $\sum_{x=0}^{\infty} P(x) = 1$

$$P(X=0) + P(X=1) + P(X=2) + P(X=3) + P(X=4) + P(X=5) + P(X=6) + P(X=7) = 1$$

$$0 + k + 2k + 2k + 3k + k^2 + 2k^2 + 7k^2 + k = 1$$

$$10k^2 + 9k = 1$$

$$10k^2 + 9k - 1 = 0 \qquad -10 \rightarrow -1 \times 10$$

$$10k^2 + 10k - 1k - 1 = 0$$

$$10k(k+1) - 1(k+1) = 0$$

$$(10k-1)(k+1) = 0$$

$$k = \frac{1}{10} \text{ or } -1$$

Here we should not take $k = -1$

we should consider $\boxed{k = \frac{1}{10}}$

Now find:

$$P(X \le 6) = P(X=0) + P(X=1) + P(X=2) + P(X=3) + P(X=4) + P(X=5) + P(X=6)$$

$$= 0 + k + 2k + 2k + 3k + k^2 + 2k^2$$

$$= 8k + 3k^2$$

$$= 8\left(\frac{1}{10}\right) + 3\left(\frac{1}{10}\right)^2$$

$$= \frac{8}{10} + \frac{3}{100}$$

$$= 0.8 + 0.03$$

$$\boxed{P(X \le 6) = 0.83} \quad \cancel{\frac{83}{100}}$$

(or)

$$P(X \le 6) = \frac{83}{100}$$

9(a) Let $Y = 2X+3$ if the random variable is uniformly distributed over $[-1, 2]$ determine $f_Y(y)$.

Ans:

Given $y = 2x+3$

given interval $[\underset{a}{-1}, \underset{b}{2}]$

If random variable is uniformly distributed

$$f_X(x) = \begin{cases} \frac{1}{b-a}, & a < x < b \\ 0, & \text{otherwise} \end{cases}$$

$$= \begin{cases} \frac{1}{2-(-1)}, & -1 < x < 2 \\ 0, & \text{otherwise} \end{cases}$$

$$f_X(x) = \begin{cases} \frac{1}{3}, & -1 < x < 2 \\ 0, & \text{otherwise.} \end{cases}$$

The tranformation $y = 2x + 3$

$$2x = y-3$$

$$x = \frac{y-3}{2}$$

It canbe written as

$$x = \frac{y-3}{2}$$

diff w.r.t 'y'

$$\frac{dx}{dy} = \frac{1}{2} \frac{d}{dy}[y-3]$$

$$= \frac{1}{2}[1-0]$$

$$\frac{dx}{dy} = \frac{1}{2} \Rightarrow \boxed{\left|\frac{dx}{dy}\right| = \frac{1}{2}}$$

We know that

$$f_y(y) = f_x(x) \left| \frac{dx}{dy} \right|$$

$$= \frac{1}{3} \times \frac{1}{2}$$

$$\boxed{f_y(y) = \frac{1}{6}}$$

9(b) Explain about the characterstic function and state its properties.

Ans: <u>Characteristic function</u> :-

Consider a random variable x with a p.d.f $f_x(x)$, then the expected value of the function $e^{j\omega x}$ is called characteristic function. It is expressed as

$$\phi_x(\omega) = E[e^{j\omega x}]$$

It is a function of a real variable, $-\infty < \omega < \infty$ where j is an imaginary operator.

i.e $\phi_x(\omega) = E[e^{j\omega x}] = \sum_i e^{j\omega x_i} p(x_i)$ (for discrete)

$$\phi_x(\omega) = \int_{-\infty}^{\infty} e^{j\omega x} f_x(x)\, dx \quad \text{(for continuous)}$$

Properties of charaterstic function:

1. The characterstic function is Unity at $w=0$ and given by

$$\phi_X(w)\Big|_{w=0} = \phi_X(0) = 1$$

2. The maximum amplitude of the Characteristic function is Unity at $w=0$ i.e $|\phi_X(w)| \leq \phi_X(0)$ (or) $|\phi_X(w)| \leq 1$

3. $\phi_X(w)$ is a continuous function of w in the range $-\infty < w < \infty$.

4. $\phi_X(-w)$ and $\phi_X(w)$ are conjugate functions

5. If $\phi_X(w)$ is a charalterstic function of a random variable X, then the Characterstic function of $Y = aX + b$ is given by

6.

10. Stale and explain the properties of marginal distribution functions, Conditional distribution and density function.

Ans:

Marginal distribution function!

The distribution function $F_{XY}(x,y)$ of a single random variable (i.e either x or y) which is obtained by adjusting one of its value to infinity is known as marginal distribution function.

The marginal distribution function of x and y are given by

$$F_x(x) = F_{xy}(x,\infty) = P(X \le x) = P(X \le x, Y < \infty)$$

$$F_y(y) = F_{xy}(\infty, y) = P(Y \le y) = P(X \le \infty, Y \le y)$$

i.e $$F_x(x) = \sum_y P(X=x, Y=y)$$

$$F_y(y) = \sum_x P(X=x, Y=y)$$

For a continuous bivariate random variables x and y the marginal distribution functions are given by

$$F_x(x) = \int_{-\infty}^{x} \int_{-\infty}^{\infty} f_{xy}(x,y)\, dx\, dy$$

$$F_y(y) = \int_{-\infty}^{y} \int_{-\infty}^{\infty} f_{xy}(x,y)\, dx\, dy$$

Marginal density function:

Marginal density function are the density function of individual random variables x & y. they are given as

$$f_x(x) = \frac{d}{dx}[F_x(x)]$$

$$f_y(y) = \frac{d}{dy}[F_y(y)]$$

where $f_x(x)$ and $f_y(y)$ are the marginal density function of x and y respectively

It can also be expressed in terms of joint density functions

$$f_x(x) = \int_{-\infty}^{\infty} f_{xy}(x,y)\, dy$$

$$f_y(y) = \int_{-\infty}^{\infty} f_{xy}(x,y)\, dx.$$

Properties of Conditional distribution function:

1. $F_X(-\infty/B) = 0$

2. $F_X(\infty/B) = 1$

3. $0 \le F_X(x/B) \le 1$

4. $F_X(x_1/B) \le F_X(x_2/B)$ if $x_1 < x_2$

5. $P(x_1 < X \le x_2/B) = F_X(x_2/B) - F_X(x_1/B)$

6. $F_X(x^+/B) = F_X(x/B)$

11 a) Consider two random variables X and Y with joint probability Mass function given in Table below.

	$Y=2$	$Y=4$	$Y=5$
$X=1$	$\frac{1}{12}$	$\frac{1}{24}$	$\frac{1}{24}$
$X=2$	$1/6$	$1/12$	$1/8$
$X=3$	$\frac{1}{4}$	$\frac{1}{8}$	$\frac{1}{12}$

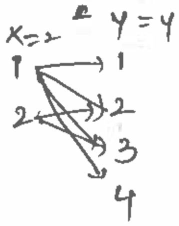
Find $P(X \le 2, Y \le 4)$, $P(Y=2/X=1)$ and check X and Y are independent

Ans: Given

	$Y=2$	$Y=4$	$Y=5$
$X=1$	$\frac{1}{12}$	$\frac{1}{24}$	$\frac{1}{24}$
$X=2$	$\frac{1}{6}$	$\frac{1}{12}$	$\frac{1}{8}$
$X=3$	$\frac{1}{4}$	$\frac{1}{8}$	$\frac{1}{12}$

We know that

$$F_{X,Y}(x,y) = P(X \le x, Y \le y)$$



(i) $P(X \le 2, Y \le 4) = P(X=1, Y=1) + P(X=1, Y=2)$

$+ P(X=1, Y=3) + P(X=1, Y=4) + P(X=2, Y=1)$

$+ P(X=2, Y=2) + P(X=2, Y=3)$

$+ P(X=2, Y=4)$

$$= 0 + \frac{1}{12} + 0 + \frac{1}{24} + \frac{1}{6} + \frac{1}{12}$$

$$= \frac{2}{12} + \frac{1+4}{24} = \frac{1}{6} + \frac{5}{24}$$

$$= \frac{4+5}{24} = \frac{\cancel{9}^{3}}{\cancel{24}_{8}}$$

$$\therefore \boxed{P(X \le 2, Y \le 4) = \frac{3}{8}}$$

(ii) $P(Y=2/X=1) = \dfrac{P(X=1, Y=2)}{P(X=1)}$

Now find $P(X=1) = \frac{1}{12} + \frac{1}{24} + \frac{1}{24}$

$$= \frac{2+1+1}{24} = \frac{4}{24} = \frac{1}{6}$$

$$P(X=1) = \frac{1}{6}$$

$$P(X=1, Y=2) = \frac{1}{12}$$

$$= \frac{\frac{1}{\cancel{12}\,2}}{\frac{1}{\cancel{6}}}$$

$$\boxed{P(Y=2/X=1) = \frac{1}{2}}$$

11) b) State and prove central limit theorem for equal distributions.

Ans: Central limit theorem :-

According to the central limit theorem (CLT), the distribution of random process which is cumulative effect of a large number of independent noise sources can be assumed to be Gaussian.

For example in communication systems, the noise is always modelled as a random variable with gaussian distribution. This is a valid assumption since the noise in a communication system is the cumulative effect of many random noise sources.

Equal distributions :-

Consider N continuous random variables X_n, $n=1,2,3,\ldots N$ have the same distribution and density functions.

Let $Y = X_1 + X_2 + X_3 + \cdots + X_N$

Also let W be normalized random variable

i-e $$W = \frac{Y-\bar{Y}}{\sigma_Y}$$

where $Y = \sum_{n=1}^{N} X_n$, $\bar{Y} = \sum_{i=1}^{N} \bar{X_n}$ and $\sigma_Y^2 = \sum_{n=1}^{N} \sigma_{X_n}^2$

So $$W = \frac{\sum_{n=1}^{N} x_n - \sum_{n=1}^{N} \bar{x}_n}{\left(\sum_{n=1}^{N} \sigma_{x_n}^2\right)^{1/2}}$$

since all random variables have the same distribution

$$\sigma_{x_n}^2 = \sigma_x^2, \left(\sum_{n=1}^{N} \sigma_x^2\right)^{1/2} = (N \sigma_x^2)^{1/2}$$
$$= \sqrt{N \sigma_x^2} = \sqrt{N} \sigma_x$$

and $\bar{x}_n = \bar{x}$

$$\therefore W = \frac{1}{\sqrt{N}\sigma_x} \sum_{n=1}^{N} (x_n - \bar{x})$$

Then w is a gaussian random variables.

12) a) Explain about second order and (or) wide sense stationary process :

Ans:- <u>Second order or wide Sense Stationary process</u>:-

If the second order probability density function of random process is independent of time (i.e does not change with time) then it is called as second order stationary process

i.e $f_x(x_1, x_2; t_1, t_2) = f_x(x_1, x_2; t_1+\tau, t_2+\tau)$

where, τ is a real number

Since second order probability density function determines the first order density function, a second order stationary process consists of first order stationary process.

The auto correlation function of the random process x(t) for t_1 and t_2 is given as,

$$R_{XX}(t_1, t_2) = E\left[x(t_1)\, x(t_2)\right]$$

If the auto correlation function of the random process is the function of time difference ($\tau = t_2 - t_1$) but not absolute time, then it is called as second order (or) wide sense stationary process (WSS)

$$\text{i.e } R_{XX}(t_1, t_1+\tau) = E\left[x(t_1)\, x(t_1+\tau)\right]$$

$$\therefore R_{XX}(t_1, t_1+\tau) = R_{XX}(\tau)$$

A random process is said to be wide sense stationary or weak stationary if it satisfies the following conditions,

1. The mean of the random process is constant

$$\text{i.e } E[x(t)] = \bar{x} = \text{Constant.}$$

2. Its auto correlation function depends only on τ (i.e $t_2 - t_1$) but not on t.

$$\text{i.e } E[x(t)\, x(t+\tau)] = R_{XX}(\tau)$$

12) b) Derive the Relationship between cross-power density spectrum and cross-correlation function.

Ans!

Let $X_T(t)$ and $Y_T(t)$ are ensemble members of the process $X(t)$ and $Y(t)$ respectively. The fourier transform of $X_T(t)$ is given by,

$$X_T(\omega) = \int_{-T}^{T} X(t)\, e^{-j\omega t}\, dt \rightarrow ①$$

The fourier transform of $Y_T(t)$ is given by

$$Y_T(\omega) = \int_{-T}^{T} Y(t_1)\, e^{-j\omega t_1} dt_1 \rightarrow ②$$

The fourier transform of $x_T^*(t)$ is given by

$$\Rightarrow X_T^*(\omega) = \int_{-T}^{T} X(t)\, e^{j\omega t}\, dt \rightarrow ③$$

Multiplying equation (3) and equation (2)

$$\therefore X_T^*(\omega) \cdot Y_T(\omega) = \int_{-T}^{T} X(t)\, e^{j\omega t} dt \int_{-T}^{T} Y(t_1)\, e^{-j\omega t_1} dt_1 \rightarrow ④$$

But we have

$$S_{XY}(\omega) = \underset{T\to\infty}{Lt} \frac{E[X_T^*(\omega)\, Y_T(\omega)]}{2T} \rightarrow ⑤$$

On substituting equation (4) in equation (5), we get

$$S_{XY}(\omega) = \underset{T\to\infty}{Lt}\; E\left[\int_{-T}^{T} X(t)\, e^{j\omega t} dt \int_{-T}^{T} Y(t_1)\, e^{-j\omega t_1} dt_1\right]$$

$$\therefore S_{XY}(\omega) = \underset{T\to\infty}{Lt} \frac{1}{2T} \int_{-T}^{T}\int_{-T}^{T} R_{XY}(t,t_1)\, e^{-j\omega(t-t_1)}\, dt\,dt_1 \rightarrow ⑥$$

Now applying inverse transform on both side, we get,

$$\frac{1}{2T}\int_{-\infty}^{\infty} S_{xy}(w)\, e^{jw\tau} dw = \frac{1}{2\pi}\int_{-\infty}^{\infty} \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{xy}(t,t_1)\, e^{-jw(t_1-t)}\, dt\, dt_1\; e^{jw\tau} d\tau$$

$$= \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{xy}(t,t_1)\, \frac{1}{2\pi}\int_{-\infty}^{\infty} e^{jw(\tau-t_1+t)}\, dw\, dt\, dt_1$$

$$\Rightarrow \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{xy}(t,t_1)\, \delta(t_1-\tau-t)\, dt_1\, dt \quad \left[\because \frac{1}{2\pi}\int_{-\infty}^{\infty} e^{jw(\tau-t_1+t)} dw = \delta(t_1-\tau-t)\right]$$

$$= \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T} R_{xy}(t,t+\tau)\, dt$$

∴ The relation becomes

$$\frac{1}{2\pi}\int_{-\infty}^{\infty} S_{xy}(w)\, e^{jw\tau} dw = \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T} R_{xy}(t,t+\tau)\, dt$$

The above relation is valid only for $-T < t+\tau < T$.

(OR)

13) a) Explain how random processes are classified with neat sketches.

Ans:- <u>Deffinition of Random Process</u> :-

A random variable which is a function of sample space and time is called random process $\{x(t,s)\}$.

It can be represented as a time function over the entire time i.e $\boxed{x(t,s) = x(t)}$.

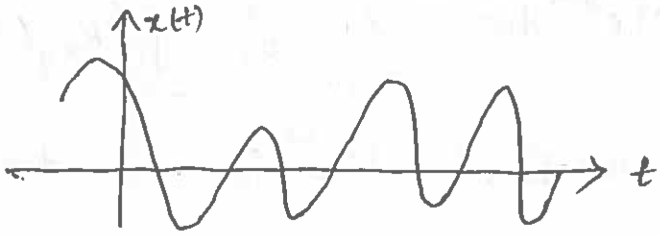
Classification of Random process :-

Random processes are mainly classified into four types based on time 't' and amplitude of random variable 'X'.

1. Continuous Random process (CRP)
2. Discrete Random process (DRP)
3. Continuous Random Sequence (CRS)
4. Discrete Random Sequence (DRS)

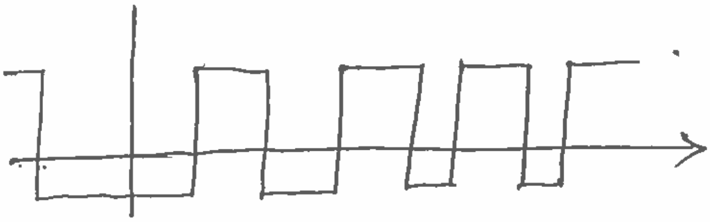
1. Continuous Random process :-

A random process is said to be continuous if random variable X and time are continuous over the entire time



2. Discrete Random process :-

In discrete random process, the random variable X has only discrete value while time 't' is continuous

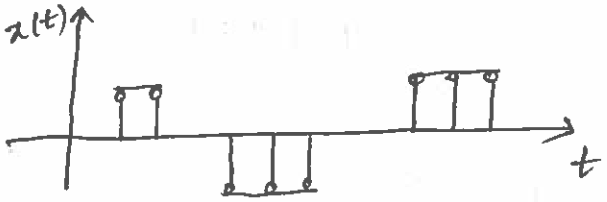


c) Continuous random process :-

In Continuous random sequence, random variable X is continuous but time t has discrete values.

d) Discrete Random Sequence :-

In discrete random sequence both random variables X and time 't' are discrete.



13) b) Derive the relationship between power spectral density and auto correlation.

Ans: Relationship between power spectral density and auto correlation :-

Let $X(t)$ be an ensemble of random process $X(t)$.

Let us define $X_T(t)$ as

$$X_T(t) = \begin{cases} x(t) & \text{for } -T \leq t \leq T \\ 0 & \text{, otherwise} \end{cases}$$

The fourier transform of $x_T(t)$ is given by

$$X_T(j\omega) = \int_{-T}^{T} x_T(t)\, e^{-j\omega t}\, dt$$

Using parseval's theorem, we can write

$$\int_{-T}^{T} x^2(t)\,dt = \frac{1}{2\pi}\int_{-\infty}^{\infty} |X_T(j\omega)|^2\,d\omega$$

The average power is given by

$$P(T) = \frac{1}{2T}\int_{-T}^{T} x^2(t)\,dt = \frac{1}{2\pi}\int_{-\infty}^{\infty} \frac{|X_T(j\omega)^2|}{2T}\,d\omega$$

To find the average power of the random process, we take the expected value with T tending to infinity in the above equation.

$$P_{XX} = \mathop{Lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T} E[X^2(t)]\,dt$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} \mathop{Lt}_{T\to\infty} \frac{E\left[|X_T(j\omega)|^2\right]}{2T}\,d\omega$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)$$

$$S_{XX}(\omega) = \mathop{Lt}_{T\to\infty} E\left[\frac{|X_T(j\omega)|^2}{2T}\right]$$

We know

$$X_T(j\omega) = \int_{-T}^{T} x_T(t)\,e^{-j\omega t}\,dt$$

Using the above equation, we can obtain

$$S_{XX}(\omega) = \mathop{Lt}_{T\to\infty} \cdot E\left[\frac{1}{2T}\int_{-T}^{T} x(t_1)\,e^{j\omega t_1}\,dt_1 \int_{-T}^{T} x(t_2)\,e^{-j\omega t_2}\,dt_2\right]$$

$$= \mathop{Lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} E\left[x(t_1)\,x(t_2)\right] e^{-j\omega(t_2-t_1)}\,dt_2\,dt_1$$

We know $E[x(t_1)\,x(t_2)] = R_{XX}(t_1, t_2)$

$$S_{XX}(\omega) = \mathop{Lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\,e^{-j\omega(t_2-t_1)}\,dt_2\,dt_1$$

Now the inverse fourier transform of $S_{XX}(\omega)$ is

$$F^{-1}[S_{XX}(\omega)] = \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)\, e^{j\omega\tau}\, d\omega$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} \lim_{T\to\infty} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, e^{-j\omega(t_2-t_1)}\, e^{j\omega\tau}\, dt_2\, dt_1\, d\omega$$

$$= \mathop{lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, \frac{1}{2\pi}\int_{-\infty}^{\infty} e^{j\omega(t_2-t_1-\tau)}\, d\omega\, dt_2\, dt_1$$

$$= \mathop{lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, \frac{1}{2\pi}\left[2\pi\, \delta(t_2-t_1-\tau)\, d\omega\, dt_2\, dt_1\right.$$

$$= \mathop{lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, \delta(t_2-t_1-\tau)\, dt_2\, dt_1$$

$$= \mathop{lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T} R_{XX}(t_1, t_1+\tau)\, dt_1$$

$$= \mathop{lt}_{T\to\infty} \frac{1}{2T}\int_{-T}^{T} R_{XX}(t, t+\tau)\, dt$$

$$= A\left[R_{XX}(t, t+\tau)\right]$$

which is the time average of the process autocorrelation function.

The above equation says that the inverse fourier transform of the power density spectrum is the time average of the auto correlation function of the process. Therefore, we can write

$$S_{XX}(\omega) = \int_{-\infty}^{\infty} A\left[R_{XX}(t, t+\tau)\right] e^{-j\omega\tau}\, d\tau$$

for a wide-sense stationary process

$$A\left[R_{XX}(t, t+\tau)\right] = R_{XX}(\tau)$$

Therefore

$$S_{XX}(w) = \int_{-\infty}^{\infty} R_{XX}(\tau)\, e^{-jw\tau}\, d\tau \rightarrow ①$$

and

$$R_{XX}(\tau) = \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(w)\, e^{jw\tau}\, dw \rightarrow ②$$

i-e power spectral density is fourier transform of auto correlation function. and eqn ① & ② is also known as Wiener - Khintchine relations.

14) a) A random processes $X(t) = A\sin(wt+\theta)$, where A, w are Constants and θ is Uniformly distributed random variable on the interval $(-\pi, \pi)$. find average power?

Ans:-

Given $X(t) = A\sin(wt+\theta)$

θ is uniform distributed on the interval $(\underset{a}{-\pi}, \underset{b}{\pi})$

$$f_\theta(\theta) = \begin{cases} \frac{1}{b-a}, & a \le \theta \le b \\ 0, & \text{otherwise} \end{cases}$$

$$= \begin{cases} \frac{1}{\pi-(-\pi)}, & -\pi \le \theta \le \pi \\ 0, & \text{otherwise} \end{cases}$$

$$f_\theta(\theta) = \begin{cases} \frac{1}{2\pi}, & -\pi \le \theta \le \pi \\ 0, & \text{otherwise} \end{cases}$$

$$P_{xx}(\tau) = E[x^2(t)]$$

$$= \int_{-\infty}^{\infty} x^2(t)\, f_\theta(\theta)\, d\theta$$

$$= \int_{-\pi}^{\pi} [A \sin(\omega t+\theta)]^2 \cdot \frac{1}{2\pi}\, d\theta$$

$$= \frac{1}{2\pi} \int_{-\pi}^{\pi} A^2 \sin^2(\omega t+\theta)\, d\theta$$

$$= \frac{A^2}{2\pi} \int_{-\pi}^{\pi} \left[\frac{1-\cos 2(\omega t+\theta)}{2}\right] d\theta$$

$$= \frac{A^2}{2\pi} \times \frac{1}{2} \int_{-\pi}^{\pi} [1-\cos(2\omega t+2\theta)]\, d\theta$$

$$= \frac{A^2}{4\pi} \left[\theta - \frac{\sin(2\omega t+2\theta)}{2} \right]_{-\pi}^{\pi}$$

$$= \frac{A^2}{4\pi} \left[[\theta]_{-\pi}^{\pi} - \frac{1}{2}[\sin(2\omega t+2\theta)]_{-\pi}^{\pi} \right]$$

$$\Rightarrow \frac{A^2}{4\pi} \left[[\pi-(-\pi)] - \frac{1}{2}\left[\sin(2\omega t+2\pi) - \sin(2\omega t-2\pi) \right] \right]$$

$$= \frac{A^2}{4\pi} \left[2\pi - \frac{1}{2}\left[\sin(2\pi+2\omega t) + \sin(2\pi-2\omega t) \right] \right]$$

$$= \frac{A^2}{4\pi} \left[2\pi - \frac{1}{2}\left[\sin 2\omega t - \sin 2\omega t \right] \right]$$

$$= \frac{A^2}{4\pi} \cancel{\times} \left[2\pi - \frac{1}{2}(0) \right]$$

$$= \frac{A^2}{\underset{2}{\cancel{4\pi}}} \times \cancel{2\pi} = \frac{A^2}{2}.$$

$$\therefore \boxed{P_{xx}(\tau) = \frac{A^2}{2}}$$

14) b) Derive the relationship between input PSD and output PSD of an LTI System.

Ans: **Power Density Spectrum of Response :-**

Consider that a random process $X(t)$ is applied on an LTI system having a transfer function $H(\omega)$. The output response is $Y(t)$.

If the power Spectrum of the input is $S_{XX}(\omega)$, then then power spectrum of the output response is given by

$$S_{YY}(\omega) = |H(\omega)|^2 S_{XX}(\omega)$$

Proof :- Let $R_{XX}(\tau)$ be the autocorrelation of the output response $Y(t)$. then the power Spectrum of the response is the fourier transform of $R_{YY}(\tau)$.

$$\therefore S_{YY}(\omega) = F[R_{YY}(\tau)]$$

$$= \int_{-\infty}^{\infty} R_{YY}(\tau)\, e^{-j\omega\tau}\, d\tau.$$

We know that $R_{YY}(\tau) = \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} R_{XX}(\tau+\tau_1-\tau_2)\, h(\tau_1)\, h(\tau_2)\, d\tau_1\, d\tau_2$

Then $S_{YY}(\omega) = \int_{-\infty}^{\infty}\int_{-\infty}^{\infty}\int_{-\infty}^{\infty} R_{XX}(\tau+\tau_1-\tau_2)\, h(\tau_1)\, h(\tau_2)\, d\tau_1\, d\tau_2\, e^{-j\omega t}\, d\tau.$

$$= \int_{-\infty}^{\infty} h(\tau_1) \int_{-\infty}^{\infty} h(\tau_2) \int_{-\infty}^{\infty} R_{XX}(\tau+\tau_1-\tau_2)\, e^{-j\omega\tau}\, d\tau\, d\tau_2\, d\tau_1$$

$$= \int_{-\infty}^{\infty} h(\tau_1)\, e^{j\omega\tau_1} \int_{-\infty}^{\infty} h(\tau_2)\, e^{-j\omega\tau_2} \int_{-\infty}^{\infty} R_{xx}(\tau+\tau_1-\tau_2)\, e^{-j\omega t}\, e^{-j\omega\tau_1}\, e^{j\omega\tau_2}\, d\tau\, d\tau_1\, d\tau_2$$

Let $\tau+\tau_1-\tau_2 = t$, $d\tau = dt$

$$S_{yy}(\omega) = \int_{-\infty}^{\infty} h(\tau_1)\, e^{j\omega\tau_1} d\tau_1 \int_{-\infty}^{\infty} h(\tau_2)\, e^{-j\omega\tau_2} \int_{-\infty}^{\infty} R_{xx}(t)\, e^{-j\omega t}\, dt$$

We know that

$$S_{yy}(\omega) = H^{*}(\omega)\, H(\omega)\, S_{xx}(\omega) = H(-\omega)\, H(\omega)\, S_{xx}(\omega)$$

$$\boxed{S_{yy}(\omega) = |H(\omega)|^2 S_{xx}(\omega)}$$

(OR)

15) a) Explain the following (i) Noise figure (ii) Noise Sources.

Ans:

Noise Figure :-

Noise figure gives the amount of noise internally generated by the system. It is the ratio of the power density of the total noise available at the output of the network to the power density available at the output only due to the input noise source. Noise figure gives a measure of the system performance of the noise.

It is mathematically expressed as

$$F = \frac{S_{no}(\omega)}{S'_{no}(\omega)} = \frac{S'_{no}(\omega) + S''_{no}(\omega)}{S'_{no}(\omega)} = 1 + \frac{S''_{no}(\omega)}{S'_{no}(\omega)}$$

where $S_{no}(\omega)$ = the total noise power spectral density at the output $S'_{no}(\omega)$ is the noise power spectral density at the output due to input noise and $S''_{no}(\omega)$ = noise power spectral density at the output due to the noise generated internally by the system. and $S_{no}(\omega) = S'_{no}(\omega)$.

then $F = 1$

Note: If $F > 1$, the system is said to be a noisy system.

There range of F is $1 < F < \infty$. As F increases, the system becomes noisy.

(ii) Noise Sources:-

There are two types of noise sources.

1. External noise
2. Internal noise.

External Noises:-

Noise whose sources are external to the receiver is called external noise. Most external noise is added into the desired signal is communication channels.

In external noise having (i) Atmospheric Noise

(ii) Extraterrestrial Noise

(iii) Industrial Noise

Internal Noise :-

The noise Created with in a device or a system is called internal noise.

→ Internal noise generated by any of the active (or) passive devices found in systems. This noise is also called function noise.

Internal noise having → (i) Shot noise
(ii) Transit - time noise
(iii) Flicker Noise
(iv) Thermal Noise.

15) b) Derive the expression for average cross power between two random process x(t) and y(t)

Ans:-

Cross power density Spectrum:

given two random process x(t) and y(t).

$$S_{xy}(\omega) = \int_{-\infty}^{\infty} R_{xy}(\tau) e^{-j\omega\tau} d\tau$$

$$S_{yx}(\omega) = \int_{-\infty}^{\infty} R_{yx}(\tau) e^{-j\omega\tau} d\tau$$

$$R_{yx}(\tau) = \frac{1}{2\pi} \int_{-\infty}^{\infty} S_{yx}(\omega) e^{j\omega\tau} d\omega$$

$$R_{xy}(\tau) = \frac{1}{2\pi} \int_{-\infty}^{\infty} S_{xy}(\omega) e^{j\omega\tau} d\omega.$$

Consider two real random process $x(t)$ and $y(t)$. If $x(t)$ and $y(t)$ are jointly WSS random process then the Cross power density spectrum is defined as F.T of the cross Correlation function of $x(t)$ and $y(t)$

$$S_{xy}(\omega) = \underset{T\to\infty}{Lt} \frac{E\left[|X_T^*(\omega)\, Y_T(\omega)\right]}{2T}$$

$$S_{yx}(\omega) = \underset{T\to\infty}{Lt} \frac{E\left[|X_T(\omega)\, Y_T^*(\omega)\right]}{2T}$$

Average Cross power :-

The average cross power of a wide sense Stationary random process. $x(t)$ is the any power is the Cross correlation function at $\tau=0$

$$P_{xy} = R_{xy}(\tau)\,|_{\tau=0}$$

$$= E\left[x(t)\, y(t+\tau)\right]/_{\tau=0}$$

$$\boxed{P_{xy} = \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{xy}(\omega)\,d\omega}$$

(or) Time average of 2nd moment.

$$P_{xy} = A\left[E[x(t)\, y(t)]\right]$$

$$= \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T} E\left[x(t)\, y(t)\right] dt$$

$$P_{xy} = \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T} R_{xx}(0)\, dt$$

——— x ———

NSRIT

Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	Program	CSE, CSE (AI & ML) & CSE (DS)			Academic Year	2020 - 2021
Course Code	20CS304	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	Object Oriented Programming through C++						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Differentiate formatted and unformatted console I/O operations.	20CS304.1	L1
2	Define Class & Object.	20CS304.2	L1
3	List any four types of Inheritance.	20CS304.3	L1
4	Define Pointers.	20CS304.4	L1
5	What are the differences between Vectors & Lists	20CS304.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Write differences between difference between Procedural Oriented Programming and object-oriented programming.	6M	20CS304.1	L1
6 (b)	Explain the advantages of object-oriented programming.	6M	20CS304.1	L2
	OR			
7 (a)	Explain the different types of Data types with example.	7M	20CS304.1	L2
7 (b)	What is code reusability? Explain different C++ features that enable reusability.	5M	20CS304.1	L1
8 (a)	With an example explain the syntax for declaring objects.	4M	20CS304.2	L2
8 (b)	What is dynamic binding? How it is different from static binding? List some advantages of dynamic binding over static binding.	8M	20CS304.2	L1
	OR			
9 (a)	Explain Friend Function.	4M	20CS304.2	L2
9 (b)	Define inline function. Write a C++ program for finding the area of a triangle using inline functions.	8M	20CS304.2	L1
10 (a)	What is a constructor? Explain with an example.	6M	20CS304.3	L1
10 (b)	Write a C++ program to find the area of circle, rectangle and triangle using function overloading.	6M	20CS304.3	L3
	OR			
11 (a)	What is inheritance? Present the advantages and disadvantages of inheritance.	6M	20CS304.3	L1
11 (b)	Write a C++ Program to overload + operator to add two matrices.	6M	20CS304.3	L3
12 (a)	Explain the role of this pointer in C++ with a programming example.	5M	20CS304.4	L2
12 (b)	What is meant by late binding? How is it implemented in C++?	7M	20CS304.4	L3
	OR			
13 (a)	What is a virtual destructor? Explain with an example.	6M	20CS304.4	L2
13 (b)	Write a C++ program to illustrate catching all exceptions.	6M	20CS304.4	L3
14 (a)	Discuss about STL programming model.	6M	20CS304.5	L2
14(b)	Write a function template for finding the minimum value in an array.	6M	20CS304.5	L3
	OR			
15 (a)	Define template. What is the need for templates in programming? Write C++ code that declares a Template class.	5M	20CS304.5	L2
15 (b)	Write a C++ program that fills a vector with random numbers.	7M	20CS304.5	L3

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SCHEME OF VALUATION
&
ANSWER KEY

Degree	B. Tech. (U. G.)	**Program**	CSE/CSM/CSD	**Test**	**END EXAM**	**Academic Year**	2021 - 2022
Course Code	**20CS304**	**Test Duration**	180 Min.	**Max. Marks**	70	**Semester**	III
Course	**Object Oriented Programming through c++**						

Assessment Pattern
R (L1): U (L2): Apply (L3): Analyze (L4): E (L5): C (L6)

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)		Learning Outcome (s)	DoK
1	**Differentiate formatted and unformatted console I/O operations.**	**2M**	**20CS304.1**	**L1**

Unformatted input /output is the most basic form of input/output.

Unformatted I/O transfers the internal binary representation of the data directly between memory and the file.

Formatted output converts the internal binary representation of the data to Ascii characters which are written to the output file.

Formatted input reads characters from the input file and converts them to internal form.

Unformatted I/O operations:cout,cin,put(),get(),getline(),write()

Formatted I/O operations :ios class functions and flags,manipulators,user-defined output functions

2	Define class and object	2M	**20CS304.2**	**L1**

- A Class is a user defined data-type which has data members and member functions.
- Data members are the data variables and member functions are the functions used to manipulate

these variables and together these data members and member functions defines the properties and behavior of the objects in a Class.

An **Object** is an instance of a Class. When a class is defined, no memory is allocated but when it is instantiated (i.e. an object is created) memory is allocated.

3 **List any four types of inheritance** **2m** **20CS304.3** **L1**

Single inheritance
Multiple inheritance
Hierarchical Inheritance
Hybrid Inheritance
Multilevel inheritance

4 Define Pointers ? **2m** **20CS304.4** **L1**

A pointer is **a variable that stores the memory address of an object**. Pointers are used extensively in both C and C++ for three main purposes: to allocate new objects on the heap, to pass functions to other functions. to iterate over elements in arrays or other data structures.

5 **What are the difference between vectors & lists ? 20CS304.5** **L1** **2m**

Vector	List
It has contiguous memory.	While it has non-contiguous memory.
It is synchronized.	While it is not synchronized.
Vector may have a default size.	List does not have default size.
In vector, each element only requires the space for itself only.	In list, each element requires extra space for the node which holds the element, including pointers to the next and previous elements in the list.
Insertion at the end requires constant time but insertion elsewhere is costly.	Insertion is cheap no matter where in the list it occurs.
Vector is thread safe.	List is not thread safe.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Learning Outcome (s)	DoK
6 (a)	Write difference between procedural oriented programming and object –oriented programming ? 6m procedure oriented-3m object oriented-3m	20CS304.1	L1

Procedural Oriented Programming	**Object-Oriented Programming**
In procedural programming, the program is divided into small parts called ***functions***.	In object-oriented programming, the program is divided into small parts called ***objects***.
Procedural programming follows a ***top-down approach***.	Object-oriented programming follows a ***bottom-up approach***.
There is no access specifier in procedural programming.	Object-oriented programming has access specifiers like private, public, protected, etc.
Adding new data and functions is not easy.	Adding new data and function is easy.
Procedural programming does not have any proper way of hiding data so it is ***less secure***.	Object-oriented programming provides data hiding so it is ***more secure***.

6 (b)	Explain the advantages of object oriented programming 6M	20ESX02.1	L2

- We can build the programs from standard working modules that communicate with one another, rather than having to start writing the code from scratch which leads to saving of development time and higher productivity,
- OOP language allows to break the program into the bit-sized problems that can be solved easily (one object at a time).
- The new technology promises greater programmer productivity, better quality of software and lesser maintenance cost.
- OOP systems can be easily upgraded from small to large systems.
- It is possible that multiple instances of objects co-exist without any interference,
- It is very easy to partition the work in a project based on objects.

OR

7 (a)	Explain the different types of data types with example. 7M Datatypes-4m examples-3m	20CS304.1	L2

Data types in C++ is mainly divided into three types

1. **Primitive Data Types**: These data types are built-in or predefined data types and can be used directly by the user to declare variables. example: int, char , float, bool etc. Primitive data types available in C++ are:

- Integer
- Character
- Boolean
- Floating Point
- Double Floating Point
- Valueless or Void
- Wide Character

2**Derived Data Types:** The data-types that are derived from the primitive or built-in datatypes are referred to as Derived Data Types. These can be of four types namely:

- Function
- Array
- Pointer
- Reference

3 **Abstract or User-Defined Data Types**: These data types are defined by user itself. Like, defining a class in C++ or a structure. C++ provides the following user-defined datatypes:

- Class
- Structure
- Union
- Enumeration
- Typedef defined DataType

```
using namespace std;

int main()
{
   cout << "Size of char : " << sizeof(char)
     << " byte" << endl;
   cout << "Size of int : " << sizeof(int)
     << " bytes" << endl;
   cout << "Size of short int : " << sizeof(short int)
     << " bytes" << endl;
   cout << "Size of long int : " << sizeof(long int)
     << " bytes" << endl;
   cout << "Size of signed long int : "
<<sizeof(signedlongint)
     << " bytes" << endl;
   cout << "Size of unsigned long int : " <<
sizeof(unsigned long int)
     << " bytes" << endl;
   cout << "Size of float : " << sizeof(float)
     << " bytes" <<endl;
   cout << "Size of double : " << sizeof(double)
     << " bytes" << endl;
   cout << "Size of wchar_t : " << sizeof(wchar_t)
     << " bytes" <<endl;

   return 0;
```

```
}
```

7 (b) What is code reusability ?explain different c++ features that enable reusability? -5m **20ESX02.3** **L3**

- C++ strongly supports the concept of reusability. The C++ classes can be reused in several ways. Once a class has been written and tested, it can be adapted by another programmer to suit their requirements. This is basically done by creating new classes, reusing the properties of the existing ones.
- software re-usability is primary attribute of software quality. C++ strongly supports the concept of reusability. C++ features such as classes, virtual function, and templates allow designs to be expressed so that re-use is made easier there are many advantage of reusability. They can be applied to reduce cost, effort and time of software development. It also increases the productivity, portability and reliability of the software product

8 (a) With an example explain the syntax for declaring objects syntax:2m example 2m 4M 20CS304.2 **L2**

An **Object** is an instance of a Class. When a class is defined, no memory is allocated but when it is instantiated (i.e. an object is created) memory is allocated.

Syntax:
ClassName ObjectName;

In C++, an object is created from a class. We have already created the class named MyClass, so now we can use this to create objects.

To create an object of MyClass, specify the class name, followed by the object name.

To access the class attributes (myNum and myString), use the dot syntax (.) on the object:

```
class MyClass {       // The class
 public:             // Access specifier
   int myNum;        // Attribute (int variable)
   string myString;  // Attribute (string variable)
};


int main() {
 MyClass myObj;  // Create an object of MyClass


 // Access attributes and set values
 myObj.myNum = 15;
 myObj.myString = "Some text";
```

```
 // Print attribute values
 cout << myObj.myNum << "\n";
 cout << myObj.myString;
 return 0;
}
```

8 (b)	**What is Dynamic binding ?how it is different from static binding?list some advantages of dynamic binding over static binding? 8m** **DYNAMIC BINDING-2M DIFFERENCE-2M ,Advantages-4m**	20CS304.2	L1

Dynamic binding refers to linking a procedure call to code that will execute only once. The code associated with the procedure is not known until the program is executed, which is also known as late binding.

1. Static binding happens when all information needed to call a function is available at the compile-time. Dynamic binding happens when the compiler cannot determine all information needed for a function call at compile-time.

Advantages:

1. Static Binding execution is more efficient & faster than Dynamic. This Binding compiler knows that these types of methods can not be overridden.
2. In the Static Binding, the type is used while Dynamic Binding uses objects for Bindings.
3. One of the major advantages of Dynamic Binding is flexibility; due to the flexibility, a single function can handle different types of an object at runtime.
4. In Static Binding, All information needed before the compilation time, while in Dynamic Binding, no information remains available before run time.
5. Static Binding can take place using normal functions, while Dynamic Binding can be achieved using virtual functions.

OR

9 (a)	**Explain Friend Function ? Explanation-1m,syntax-2m ,example-1m 4m**	20CS304.2	L2

A friend function in C++ is defined as a function that can access private, protected and public members of a class.

The friend function is declared using the friend keyword inside the body of the class.

Friend Function Syntax:

```
class className {
   ... .. ...
   friend returnType functionName(arguments);
   ... .. ...
}
```

By using the keyword, the **'friend'** compiler understands that the given function is a friend function.

We declare friend function inside the body of a class, whose private and protective data needs to be accessed, starting with the keyword friend to access the data. We use them when we need to operate between two different classes at the same time.

To declare a function as a friend of a class, precede the function prototype in the class definition with keyword **friend** as follows −

```
class Box {
  double width;

  public:
    double length;
    friend void printWidth( Box box );
    void setWidth( double wid );
};
```

9 (b)	**Define Inline function .write a c++ program for finding the area of a triangle using inline functions. Inline function-2m program -6m 8m**	20ESX02.2	L1

Inline function in C++ is an enhancement feature that improves the execution time and speed of the program. The main advantage of inline functions is that you can use them with C++ classes as well.

```
// calculate an area of triangle using the inline function
#include <iostream>
using namespace std;
```

```
// inline function, no need prototype
inline float triangle_area(float base, float height)
{
float area;
area = (0.5 * base * height);
return area;
}

int main(void)
{
float b, h, a;
b = 4;
h = 6;
// compiler will substitute the inline function code here.
a = triangle_area(b, h);
cout<<"Area = (0.5*base*height)"<<endl;
cout<<"where, base = 4, height = 6"<<endl;
cout<<"\nArea = "<<a<<endl;

return 0;
}
```

Output example:

```
Area = (0.5*base*height)
where, base = 4, height = 6
Area = 12
```

10 (a)	**What is a constructor ?Explain with an example** **constructor-2m example-2m 4m**	20CS304.3	**L1**

A constructor is a special type of member function of a class which initializes objects of a class. In C++, Constructor is automatically called when object(instance of class) is created. It is special member function of the class because it does not have any return type.

```
 using namespace std;

  class construct

 {

 public:
```

```
    int a, b;

  // Default Constructor

    construct()

    {

        a = 10;

        b = 20;

    }

 };

 int main()

 {

    // Default constructor called automatically

    // when the object is created

    construct c;

    cout << "a: " << c.a << endl

        << "b: " << c.b;

    return 1;

 }
```

Output:

```
a: 10

b: 20
```

10 (b)	**Write a c++ program to find the area of circle ,rectangle and triangle using function overloading. 6m**	**20CS304.3**	**L3**

```
* C++ program to find Area using Function Overloading  */

#include<iostream>
using namespace std;
int area(int);
int area(int,int);
float area(float);
float area(float,float);
int main()
{
        int s,l,b;
        float r,bs,ht;
        cout<<"Enter side of a square:";
        cin>>s;
        cout<<"Enter length and breadth of rectangle:";
        cin>>l>>b;
        cout<<"Enter radius of circle:";
        cin>>r;
        cout<<"Enter base and height of triangle:";
        cin>>bs>>ht;
        cout<<"Area of square is"<<area(s);
        cout<<"\nArea of rectangle is "<<area(l,b);
     cout<<"\nArea of circle is "<<area(r);
     cout<<"\nArea of triangle is "<<area(bs,ht);
 }
 int area(int s)
 {
    return(s*s);
 }
 int area(int l,int b)
 {
    return(l*b);
 }
 float area(float r)
 {
    return(3.14*r*r);
```

```
}
float area(float bs,float ht)
{
  return((bs*ht)/2);
}
```

OUTPUT : :

```
/*  C++ program to find Area using Function Overloading  */

Enter side of a square:2
Enter length and breadth of rectangle:3 6
Enter radius of circle:3
Enter base and height of triangle:4 4

Area of square is4
Area of rectangle is 18
Area of circle is 28.26
Area of triangle is 8
```

OR

11 (a) **What is inheritance ? present the advantages and disadvantages of inheritance? Defination -2m , advantages-2m ,Disadvantages-2m** 20CS304.3 **L1**

In C++, inheritance is a process in which one object acquires all the properties and behaviors of its parent object automatically. In such way, you can reuse, extend or modify the attributes and behaviors which are defined in other class.

In C++, the class which inherits the members of another class is called derived class and the class whose members are inherited is called base class. The derived class is the specialized class for the base class.

Advantages:

- Inheritance promotes reusability
- Inheritance allows us to inherit all the properties of base class and can access all the functionality of inherited class. It implements reusability of code.

Disadvantages:-

- Inherited functions work slower than normal function as there is indirection.
- Improper use of inheritance may lead to wrong solutions.
- Often, data members in the base class are left unused which may lead to memory wastage.
- Inheritance increases the coupling between base class and derived class. A change in base class

will affect all the child classes.

11 (b)	Write a c++ program to overload + operator to add two matrices. Program -5m output-1m 6m	20ESX02.3	L2

```
#include<iostream>
using namespace std;

class Matrix
{
        int a[3][3];
    public:
        void accept();
        void display();
        void operator +(Matrix x);
};
void Matrix::accept()
{
        cout<<"\n Enter Matrix Element (3 X 3) : \n";
        for(int i=0; i<3; i++)
        {
                for(int j=0; j<3; j++)
                {
                        cout<<" ";
                        cin>>a[i][j];
                }
        }
}
void Matrix::display()
{
        for(int i=0; i<3; i++)
        {
                cout<<" ";
                for(int j=0; j<3; j++)
                {
                        cout<<a[i][j]<<"\t";
```

```
            }
            cout<<"\n";
        }
    }
    void Matrix::operator +(Matrix x)
    {
        int mat[3][3];
        for(int i=0; i<3; i++)
        {
            for(int j=0; j<3; j++)
            {
                mat[i][j]=a[i][j]+x.a[i][j];
            }
        }
        cout<<"\n Addition of Matrix : \n\n";
        for(int i=0; i<3; i++)
        {
            cout<<" ";
            for(int j=0; j<3; j++)
            {
                cout<<mat[i][j]<<"\t";
            }
            cout<<"\n";
        }
    }
    int main()
    {
        Matrix m,n;
        m.accept();     // Accepting Rows
        n.accept();     // Accepting Columns
        cout<<"\n First Matrix : \n\n";
        m.display();  // Displaying First Matrix
        cout<<"\n Second Matrix : \n\n";
        n.display(); // Displaying Second Matrix
        m+n;       // Addition of Two Matrices. Overloaded '+' Operator
        return 0;
```

```
}
```

Output:

```
Enter matrix element(3*3):
4 5 6 1 2 3 7 8 9
Enter Matrix element(3*3):
1 2 3 4 5 6 7 8 9
First  Matrix::
                        4   5 6
                 1 2 3
                 7 8 9
Second matrix:
                    1 2 3
                   4 5 6
                   7 8 9



Addition of matrix:
                              5 7  9
                               5 7  9
                                           14 16 18
```

12 a) **Explain the role of this poiter in c++ with a programming example 5m**
This pointer -2m example-3m **20CS304.4** **L2**

The 'this' pointer is passed as a hidden argument to all nonstatic member function calls and is available as a local variable within the body of all nonstatic functions. 'this' pointer is not available in static member functions as static member functions can be called without any object (with class name).

```
using namespace std;

 /* local variable is same as a member's name */

class Test

{

private:

  int x;

public:

  void setX (int x)

  {

     // The 'this' pointer is used to retrieve the object's x

     // hidden by the local variable 'x'

     this->x = x;

  }
```

```
  void print() { cout << "x = " << x << endl; }
};


int main()
{
  Test obj;
  int x = 20;
  obj.setX(x);
  obj.print();
  return 0;
}
```

Output:

```
 x = 20
```

12(b) **What is meant by late binding ?how is it implemented in c++ ?** **7m**

late binding-2m implementation-5m **20CS304.4** **L3**

Late Binding : (Run time polymorphism) In this, the compiler adds code that identifies the kind of object at runtime then matches the call with the right function definition (Refer this for details). This can be achieved by declaring a virtual function.

```
using namespace std;

class Base
{
public:
   virtual void show() { cout<<" In Base \n"; }
};

class Derived: public Base
{
public:
   void show() { cout<<"In Derived \n"; }
};

int main(void)
{
   Base *bp = new Derived;
   bp->show();  // RUN-TIME POLYMORPHISM
   return 0;
}
```

Output:

```
In Derived
```

OR

13 **What is virtual destructor ?explain with an example? Virtual destructor-2m example-4m 6m 20CS304.4 L2**

A virtual destructor is used to free up the memory space allocated by the derived class object or instance while deleting instances of the derived class using a base class pointer object. A base or parent class destructor use the **virtual** keyword that ensures both base class and the derived class destructor will be called at run time, but it called the derived class first and then base class to release the space occupied by both destructors.

```
        #include<iostream>
using namespace std;
class Base
{
   public:
   Base() // Constructor member function.
{
   cout << "\n Constructor Base class";  // It prints first.
}
 virtual ~Base() // Define the virtual destructor function to call the Destructor Derived function.
{
   cout << "\n Destructor Base class";  /
}
};
// Inheritance concept
class Derived: public Base
{
   public:
   Derived() // Constructor function.
{
   cout << "\n Constructor Derived class" ; /* After print the Constructor Base, now it will prints. */
}
 ~Derived() // Destructor function
{
   cout << "\n Destructor Derived class"; /* The virtual Base Class? Destructor calls it before calling the Base
ss Destructor. */
}
};
int main()
{
   Base *bptr = new Derived; // A pointer object reference the Base class.
   delete bptr; // Delete the pointer object.
```

```
}
```

```
OUTPUT:
CONSTRUCTOR BASE CLASS
CONSTRUCTOR DERIVED CLASS
DESTRUCTOR DERIVED CLASS
DESTRUCTOR BASE CLASS
```

14 **Discuss about STL programming model** **6m** **EXPLANATION STL PROGROGRAMING-6M**

20CS304.5 L2

The Standard Template Library (STL) is a set of C++ template classes to provide common programming data structures and functions such as lists, stacks, arrays, etc. It is a library of container classes, algorithms, and iterators. It is a generalized library and so, its components are parameterized. A working knowledge of template classes is a prerequisite for working with STL.

STL has four components

- Algorithms
- Containers
- Functions
- Iterators

Algorithms

The header algorithm defines a collection of functions especially designed to be used on ranges of elements.They act on containers and provide means for various operations for the contents of the containers.

- Algorithm
  - Sorting
  - Searching
  - Important STL Algorithms
  - Useful Array algorithms
  - Partition Operations
- Numeric
  - valarray class

Containers

Containers or container classes store objects and data. There are in total seven standard "first-class" container classes and three container adaptor classes and only seven header files that provide access to these containers or container adaptors.

- Sequence Containers: implement data structures which can be accessed in a sequential manner.
  - vector
  - list
  - deque
  - arrays
  - forward_list(Introduced in C++11)
- Container Adaptors : provide a different interface for sequential containers.
  - queue
  - priority_queue
  - stack
- Associative Containers : implement sorted data structures that can be quickly searched (O(log n) complexity).
  - set
  - multiset
  - map

- multimap

Functions

The STL includes classes that overload the function call operator. Instances of such classes are called function objects or functors. Functors allow the working of the associated function to be customized with the help of parameters to be passed.

- Functors

Iterators

As the name suggests, iterators are used for working upon a sequence of values. They are the major feature that allow generality in STL.

- Iterators

OR

DEFINE TEMPLATE WHAT IS THE NEED FOR TEMPLATES IN PROGRAMMING?WRITE C++ CODE THAT DECLARES A TEMPLATE CLASS TEMPLATE -2M NEED-2M CODE-1M 20CS304.5 5M

A **template** is a simple and yet very powerful tool in C++. The simple idea is to pass data type as a parameter so that we don't need to write the same code for different data types. For example, a software company may need sort() for different data types. Rather than writing and maintaining the multiple codes, we can write one sort() and pass data type as a parameter.
C++ adds two new keywords to support templates: *'template'* and *'typename'*. The second keyword can always be replaced by keyword 'class'.

15 a) It **allows you to define the generic classes and generic functions and thus provides support for generic programming**. Generic programming is a technique where generic types are used as parameters in algorithms so that they can work for a variety of data types. Templates can be represented in two ways: Function templates

```
// C++ program to demonstrate the use of class templates

#include <iostream>

using namespace std;

// Class template
```

```
template <class T>
class Number {
  private:
   // Variable of type T
   T num;
 public:
   Number(T n) : num(n) {}   // constructor

   T getNum() {
      return num;
   }
 };
 int main() {
    // create object with int type
    Number<int> numberInt(7);
 // create object with double type
    Number<double> numberDouble(7.7);
 cout << "int Number = " << numberInt.getNum() << endl;
    cout << "double Number = " << numberDouble.getNum() << endl;
```

```
return 0;

}
```

Run Code

Output

```
int Number = 7

double Number = 7.7
```

15 b) **WRITE A C++ PROGRAM THAT FILLS A VECTOR WITH RANDOM NUMBERS 7M PROGRAM -6M OUTPUT-1M** **20CS304.5** **L3**

```
// C++ program to generate the vector
// with random values
#include <bits/stdc++.h>
using namespace std;

// Driver Code
int main()
{
    // Size of vector
    int size = 5;

    // Initialize the vector with
    // initial values as 0
    vector<int> V(size, 0);

    // use srand() for different outputs
    srand(time(0));

    // Generate value using generate
    // function
    generate(V.begin(), V.end(), rand);

    cout << "The elements of vector are:\n";

    // Print the values in the vector
    for (int i = 0; i < size; i++) {
        cout << V[i] << " ";
    }
```

Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	**Program**	Mechanical Engg.			**Academic Year**	2021 - 2022
Course Code	20ME305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	MANUFACTURING PROCESS						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Why is a loose piece pattern used? Give its problems.	20ME305.1	L1
2	What are the essential conditions that are to be kept in mind while designing risers?	20ME305.2	L2
3	Write any four differences between the welding and soldering.	20ME305.3	L2
4	Why do most welding failures occur in HAZ? Explain.	20ME305.4	L1
5	Differentiate between fullering and edging in forging operation	20ME305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the types of patterns with a neat sketch	6M	20ME305.1	L2
6 (b)	With help of sketch explain gating system.	6M	20ME305.1	L1
	OR			
7 (a)	Explain injection molding and Blow molding.	6M	20ME305.1	L2
7 (b)	Explain about hand molding process.	6M	20ME305.1	L1
8 (a)	Explain the construction and working principle of Cupola Furnace with a neat sketch.	6M	20ME305.2	L2
8 (b)	Explain the principle of gating and gating ratio	6M	20ME305.2	L2
	OR			
9 (a)	What are the types of centrifugal casting?	6M	20ME305.2	L2
9 (b)	Which centrifugal casting method is used for which application?	6M	20ME305.2	L2
10 (a)	How submerged arc welding process takes place.	7M	20ME305.3	L2
10 (b)	Describe its advantages and applications.	5M	20ME305.3	L2
	OR			
11 (a)	Explain the TIG systems of arc-welding give the applications of each.	6M	20ME305.3	L2
11 (b)	Explain the MIG systems of arc-welding give the applications of each.	6M	20ME305.3	L2
12 (a)	Describe the electro slag welding process	5M	20ME305.4	L2
12 (b)	Describe the electron beam welding process	7M	20ME305.4	L2
	OR			
13 (a)	Explain the method and application of friction stir welding.	6M	20ME305.4	L2
13 (b)	Explain any two destructive testing of welds.	6M	20ME305.4	L2
14 (a)	Describe the wire drawing process.	7M	20ME305.5	L1
14 (b)	Describe the tube drawing process.	5M	20ME305.5	L2
	OR			
15 (a)	Enumerate the typical applications of cold working.	4M	20ME305.5	L1
15 (b)	Explain the blanking and piercing process with a neat sketch.	8M	20ME305.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY
(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

ME-III sem
MANUFACTURING PROCESS (20ME 305)

Part-A (5*2=10)

1. **Why is loose piece pattern used? Give its problems.**
Loose piece of a pattern which is attached to the pattern during molding, remains in the mold during lift-off and is only then removed separately. It is used for undercuts if the mold joint cannot be positioned in this cross section.
2. **What are the essentials conditions that are to be kept in mind while designing Risers?**
1) The modulus of the riser should be larger than the modulus of the casting, which encourages directional solidification, insuring that feed metal will be available to counteract shrinkage in the casting throughout solidification
2) The riser should have enough volume to provide the required feed metal to the casting.

3. Write any four differences between Welding and Soldering?

S.NO	Welding	Soldering
1	metal fabricators melt the base metal.	metal fabricators heat the metal to be bonded but never melt them
2	Welding joints are the strongest, followed by soldered joints then brazed joints.	Soldering is most similar to brazing because it uses capillary action to flow the metal into the joint until it cools and hardens.
3	Welding requires about 6,500 degrees Fahrenheit, while	soldering requires about 840 degrees Fahrenheit.
4	Workpieces and the metal base are heated and melted in welding	Soldering requires no heating of the workpieces.

4. **Why do most welding failures occur in HAZ? Explain?**
The heat produced in the weld bead area causes chromium carbides to precipitate around the grain boundaries in the HAZ, causing the local chromium content to drop below 10.5%, at which point the steel loses the ability to form a passive film and is no longer stainless.
5. **Differentiate between fullering and edging in forging operation?**
the flow of material on impact of the die is the opposite of fullering. When the concave die deforms the workpiece the material flows into the die area from both sides. The process is called edging because it is usually carried out on the edges of the workpiece.

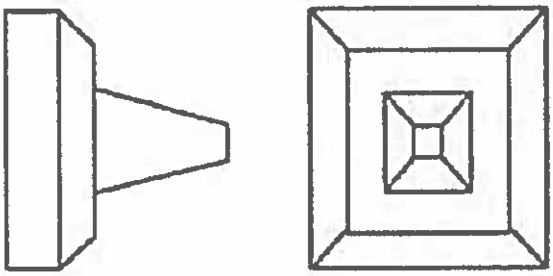
Part-B

6(a). Explain the types of patterns with neat sketch? (6M)

We use 10 different types of patterns in the casting process. Single piece pattern, two piece pattern, gated pattern, multi piece pattern, match plate pattern, skeleton pattern, sweep pattern, lose piece pattern, cope and drag pattern, shell pattern.

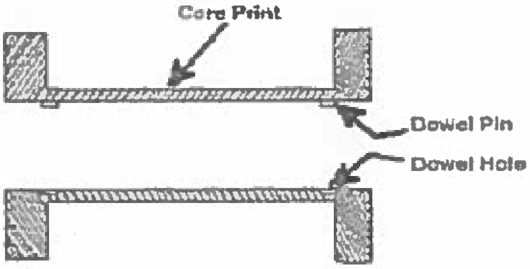
There have more details about 10 different types of patterns.

Single Piece Pattern



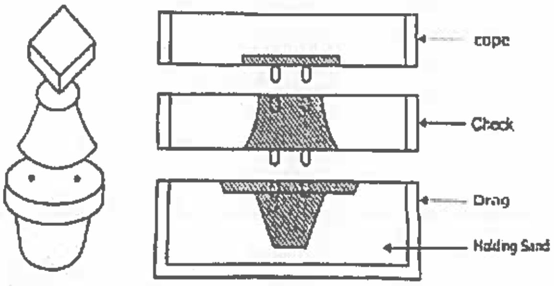
Single piece pattern, also called solid pattern is the lowest cost casting pattern. It is very suitable for simple process, and small scale production and the large casting manufacturers prefer it because this kind of casting pattern make casting process just needing simple shapes, flat surfaces like simple rectangular blocks. One flat surface is used to separate planes.

Two-Piece Pattern



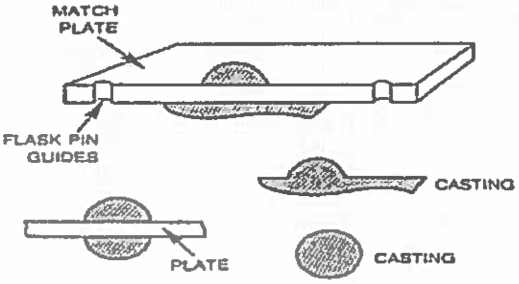
Two-piece pattern also called split piece pattern is a common casting pattern for intricate casting. This kind of pattern has parting planes which may have flat or irregular surface, and the exact position of the plane was decided by the shape of the casting. There are two pieces of the split piece pattern. One of the parts is molded in drag and another is molded in cope. And the cope part always has dowel pins. With the dowel pins, the two halves of split piece pattern can be aligned.

Multi Piece Pattern



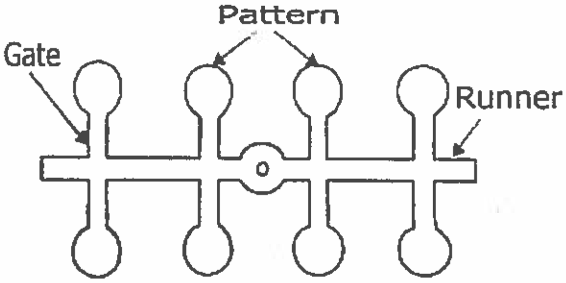
Multi piece pattern is a good solution for complex designs which is hard to make. This kind of pattern includes 3 or more pattern which helps you achieve mold making. Take the three-piece pattern as an example. The pattern is made of the top, bottom, and middle parts. The top part is cope, the bottom part drag, and the middle parts are called as checkbox.

Match Plate Pattern



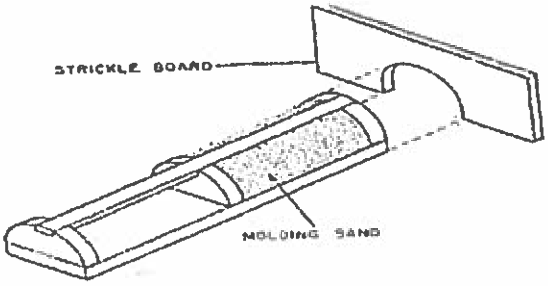
Match plate pattern has a metallic plate to divide the cope and drag areas into the opposite face of the plate. This kind of pattern nearly has no hard work and can provide high output. It is widely used in the manufacturing industry, and usually has an expensive cost, precise casting and high yield. And this kind of casting pattern is widely used in metal casting like aluminum.

Gate Pattern



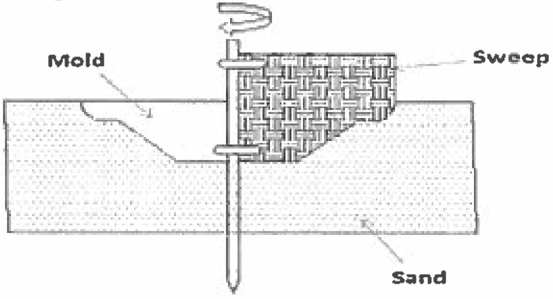
Gate pattern can consist of one or more patterns into a molding pattern. It is designed for the mold which makes multiple components at one casting process. The gates are used to combine the different patterns, and runners to create a flow way for the molten materials. When the gates and runners have already attached, the patterns are loosing. This kind of pattern is expensive, and it is usually used for small castings.

Skeleton Pattern



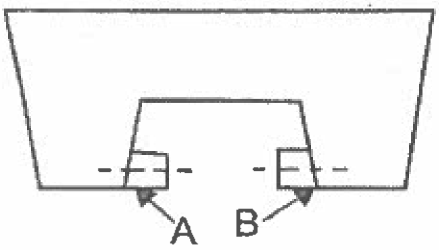
Skeleton pattern is large in size, and it is a good choice for the casting which has the simple size and shape. This kind of casting pattern is expensive and not versatile. It is not the best choice from the aspect of economic, while is very efficient in extra sand removing. If you want to use this casting pattern you should highlight the wood frames when you casting. The skeleton pattern is widely used in the industries of pit or floor welding.

Sweep Pattern



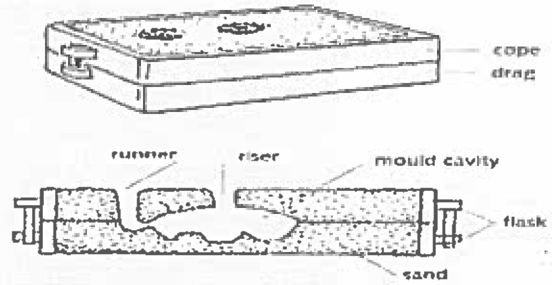
Sweep pattern uses a wooden board with proper size to rotate along one edge to shape the cavity. This kind of casting pattern creates a cavity in the vertical direction and the base of it is attached with sand, and it also creates casting in a very short time, and it has consisted of three parts: spindle, base and sweep which also called wooden board.

Loose Piece Pattern



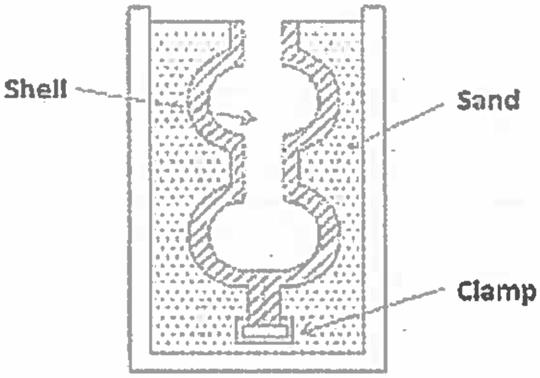
Loose piece pattern can help manufacturers remove one piece of solid pattern which is above or below the parting plane of the mold. This kind of pattern needs extra skilled labor work, so it is expensive casting pattern in castings.

Cope and Drag Pattern



Just like its name, cope and drag pattern has consisted of two separate plates, and it has two parts which can be separately molded on the molding box, and these parts create the cavity. This kind of pattern has a bit similar with the two-piece pattern and is usually used in large casting.

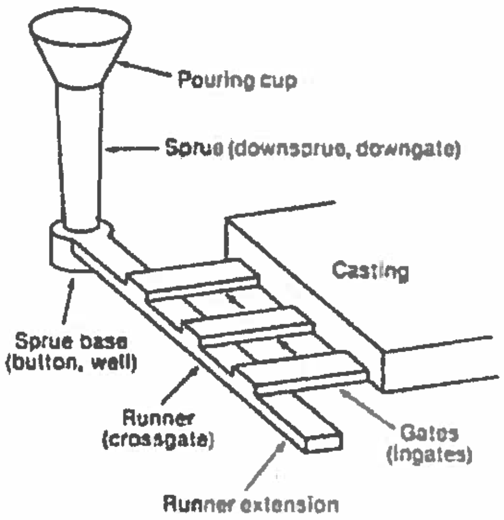
Shell Pattern



Shell pattern is a good choice to create hollow shaped structure. It parts along the center and dowels the resultant halves.

6(b). With help of sketch explain gating system? (6M)

- Elements of Gating System
  - Pouring Cup
  - Spruce
  - Spruce Well
  - Cross-gate or Runner
  - Ingate or Gates



- Pouring Cup – It is the funnel-shaped opening, made at the top of the mold. The main purpose of the pouring basin is to direct the flow of molten metal from ladle to the sprue.
- Spruce – It is a vertical passage connects the pouring basin to the runner or ingate. It is generally made tapered downward to avoid aspiration of air. The cross section of the sprue may be square, rectangular, or circular.
- Spruce Well – It is located at the base of the sprue. It arrests the free fall of molten metal through the sprue and turns it by a right angle towards the runner.
- Runner – It is a long horizontal channel which carries molten metal and distribute it to the ingates .It will ensure proper supply of molten metal to the cavity so that proper filling of the cavity takes place.
- Gate – These are small channels connecting the mould cavity and the runner.The gates used may vary in number depends on size of the casting.
  Function of Gating System
- A good gating system should help easy and complete filling of the mould cavity.
- It should fill the mould cavity with molten metal with least amount of turbulance.
- It should prevent mould erosion.
- It should establish proper temperature gradient in the casting.
- It should promote directional solidification.
- It should regulate the rate of flow of metal into the mould cavity.

(OR)

7(a). Explain Injection molding and blow molding? (6M)

Injection molding is a method to obtain molded products by injecting plastic materials molten by heat into a mold, and then cooling and solidifying them.

The method is suitable for the mass production of products with complicated shapes, and takes a large part in the area of plastic processing.

The process of injection molding is divided into 6 major steps as shown below.

1. Clamping
2. Injection
3. Dwelling
4. Cooling
5. Mold opening
6. Removal of products

The process is proceeded as shown above and products can be made successively by repeating the cycle.

Injection molding machine is divided into 2 units i.e. a clamping unit and an injection unit.

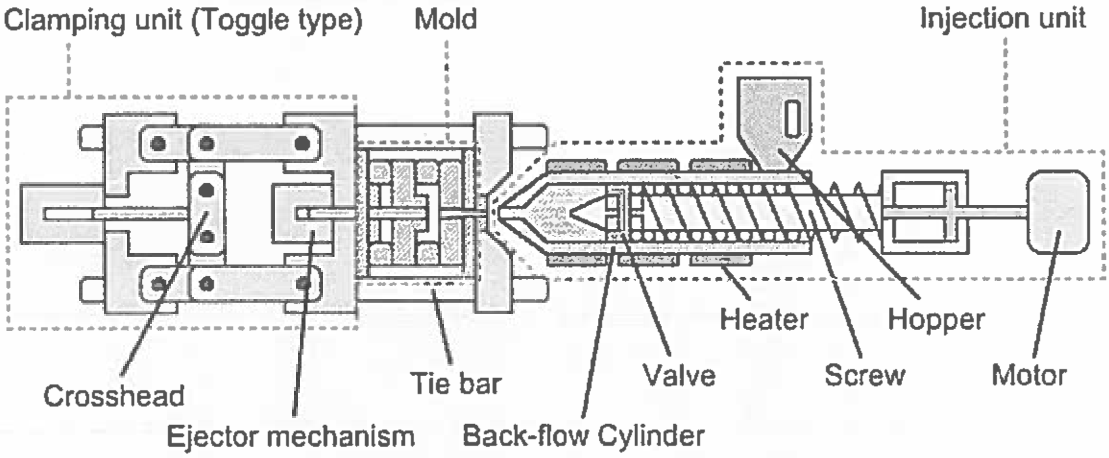
The functions of the clamping unit are opening and closing a die, and the ejection of products. There are 2 types of clamping methods, namely the toggle type shown in the figure below and the straight-hydraulic type in which a mold is directly opened and closed with a hydraulic cylinder.

The functions of the injection unit are to melt plastic by heat and then to inject molten plastic into a mold.

The screw is rotated to melt plastic introduced from the hopper and to accumulate molten plastic in front of the screw (to be called metering) . After the required amount of molten plastic is accumulated, injection process is stared.

While molten plastic is flowing in a mold, the machine controls the moving speed of the screw, or injection speed. On the other hand, it controls dwell pressure after molten plastic fills out cavities.

The position of change from speed control to pressure control is set at the point where either screw position or injection pressure reaches a certain fixed value.

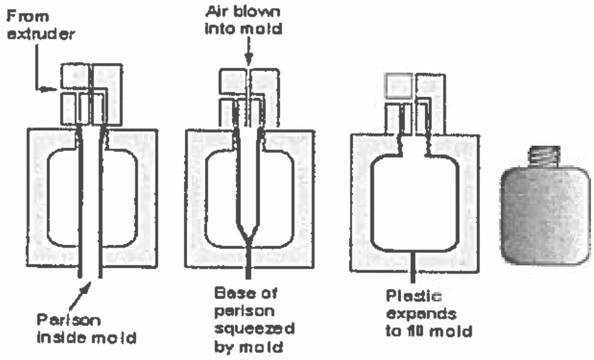


- Blow molding (or moulding) is a manufacturing process for forming and joining together hollow plastic parts. It is also used for forming glass bottles or other hollow shapes.

In general, there are three main types of blow molding: extrusion blow molding, injection blow molding, and injection stretch blow molding.

The blow molding process begins with softening plastic, by heating, and forming it into a parison or, in the case of injection and injection stretch blow molding (ISB), a preform. The parison is a tube-like piece of plastic with a hole in one end through which compressed air can pass.

The parison is then clamped into a mold and air is blown into it. The air pressure then pushes the plastic out to match the mold. Once the plastic has cooled and hardened the mold opens and the part is ejected. Water channels within the mold assist cooling.



7(b). Explain about hand molding process? (6M)

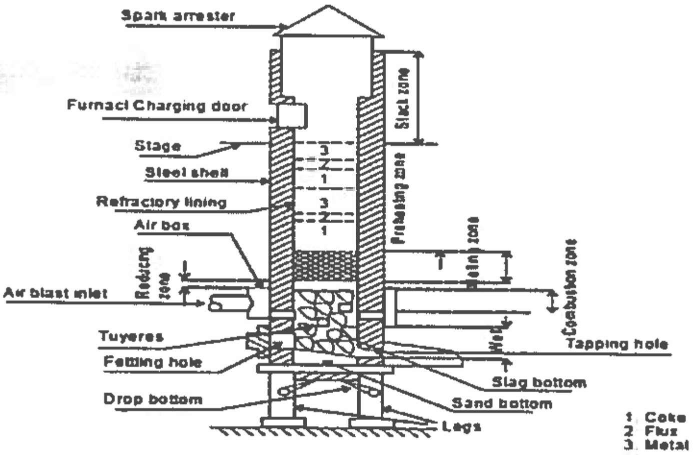
Whatever the force required for ramming and compressing of molding sand, if it is obtained from a human hand, then it is called as hand molding.

The properties of Hand Moulding method are as follows:

- It is the cheapest method of mold making operation.
- Complex shapes of the patterns will be easily used in mold making without any damages to the pattern.
- Strength and Hardness of the mold are non-uniform because of the non-uniform force applied by hand.
- The production rate is low.

The other method used for preparing the mold is Machine Molding method. It generally consists of 4 operations. By using these operations only, the mold can be prepared.

8(a). Explain the construction and working principle of cupola furnace with a neat sketch? (6M)



Working Principle of Cupola Furnace:

The Cupola furnace works on the principle where we generate heat from burning coke and when the temperature of the furnace is above the melting point of the metal then the metal is melt.

The charge introduced in the cupola consists of pig iron, scrap, casting rejection, coke, and flux. Coke is the fuel and limestone are added as a flux to remove undesirable materials like ash and dirt. The scrap consists of Steel and cast iron rejections.

The working of Cupola furnace is, Over the sand Bottom, Coke in charged extending up to a predetermined height. This serves as the coke bed within which the combustion takes place.

Cupola operation is started by igniting the coke bed at its bottom. After the Coke bed is properly Ignited, alternate charges of limestone, pig iron, and coke are charged until the level of the charging Door.

Then the air blast is turned on and combustion occurs rapidly within the coke bed. Within 5 to 10 minutes after the blast is turned on the first molten cast iron appears at the tap hole.

Usually, the first iron which comes out will be too cold to pour into sand molds. During the cupola operation, molten metal may be tracked every 10 minutes depending on the melting rate and the capacity.

All the oxygen in the air blast is consumed by the combustion, Within the combustion zone.

The chemical reaction takes place which is,

C + O2 (from the air) → CO2 + Heat

This is an exothermic reaction. The temperature in this zone varies from 1550 to 1850 degree Celsius.

Then hot gases consisting principally of Nitrogen and carbon dioxide moved upward from the combustion zone, where the temperature is 1650 degree Celsius.

The portion of the coke bed if the combustion zone is reducing zone. It is a protective zone to prevent the oxidation of the metal charge above and while dropping through it. As the hot carbon dioxide gas moves upward through the hot coke, some of it is reduced by the following reaction.

3Fe + 2CO → Fe3C + CO2

This is an endothermic reaction.

The first layer of iron above the reducing zone is the melting zone where the solid iron is converted into the molten state. A significant portion of the carbon is picked up by the metal also takes place in this zone.

The hot gas is passed upward from the reducing and melting zones into the preheating zone which includes all layers of charge above the melting zone up to the charging Door.

Since the layer of the charge is preheated by the outgoing gases which exist at the top of the cylindrical shell. this temperature is this zone is around 1090 degrees Celsius.

8(b). Explain the principle of gating and gating ratio? (6M)

A gating system should avoid sudden or right angle changes in direction. A gating system should fill the mould cavity before freezing. The metal should flow smoothly into the mould without any turbulence. A turbulence metal flow tends to form dross in the mould.

Gating ratio

Gating ratio is the ratio between the cross-sectional area of the sprue to the total cross-sectional area of the runners to the total cross-sectional area of the ingates.

The formula for the gating ratio is As: Ar: Ag.

With the Pressurized Gating System, the gating ratio is usually 1: 2: 1 or 1: 0.75: 0.5. This system is called a "*Gate control system*" because ingates control the flow of the metal.

With the Unpressurized Gating System, the gating ratio is usually 1: 2: 2 or 1: 3: 3 or 1: 1: 3. This system is called a "*Choke control system*" because the choke controls the flow of the metal.

Table of gating ratio for various of materials:

Materials	Gating ratio
Aluminum	1:2:1 1:1.2:2 1:2:4 1:3:3 1:4:4 1:6:6
Aluminum bronze	1:2.88:4.8
Brass	1:1:1 1:2:3 1.6:1.3:1
Copper	2:8:1 3:9:1
Ductile iron	1.15:1.1:1

	1.25:1.13:1 1.33:2.67:1

Gating ratio of materials

(OR)

9(a). What are the types of centrifugal casting? (6M)

Working Principle of Centrifugal Casting

In this process molten metal is poured into the spinning mold preheated to a certain temperature. The mold is placed vertically or horizontally based on the required shape of product. Once poured it is then continued to rotate about its central axis.

Due to the rotational motion of the mold; a centrifugal force is acted upon the molten metal just poured into the spinning mold. This force displaces the molten metals towards the periphery forcing them to deposit on the walls.

The molten metal is spread uniformly on to the walls of the die; thanks to the centrifugal force 100 times greater than of gravity.

As the process continues with more and more metal poured into the mold; the relatively denser element tends to deposit on towards the wall while lighter elements and slug deposit at the center.

The mold is then left to rotate till the whole mold solidify and then other light elements like slag are separated from the center.

The whole process itself leads to a reduction in defects due to slags, irregular grain structure, and trapped air. The final product have closed grain structure with improved elongation, tensile strength, and yield strength. Fig: Construction of True centrifugal casting Machine

Advantages Of Using Centrifugal Casting

The whole process of centrifugal casting does not rely on gates and risers. This results in the continuous availability of molten metal at the center during the solidification process.

Unlike conventional casting where the mold solidifies from both inside and outside; molten metal at the center ensures unidirectional solidification (Outside to Inside).

This helps get rid of defects such as blow holes, shrinkage cavity and gas pockets. The process allows for subsequent savings on initial capital (*Non-machinery cost*) and manufacturing costs.

This allows for the economic advantage which gave enough flexibility to produce various shapes and sizes of products.

The combination of deciding factors such as unidirectional solidification, solidification under pressure, and impurities displacement to the central axis; results in superior quality products with high soundness compared to other industrial processes.

Such Advantages result in increased life and endurance of the product without fracturing.

Other benefits of centrifugal casting includes:

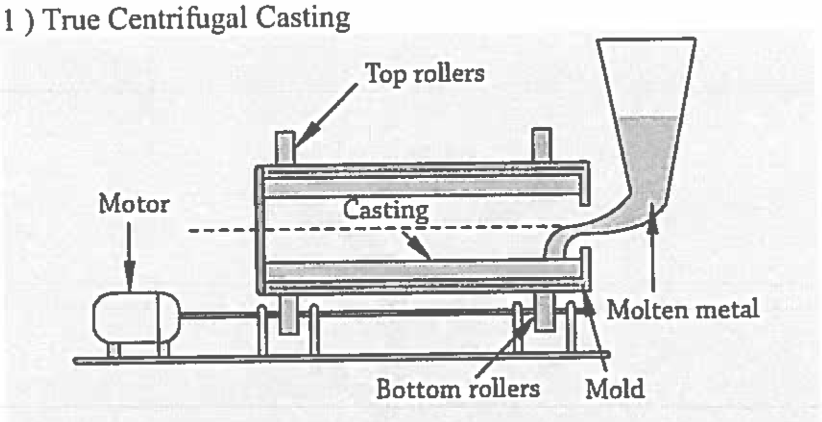
1. Impurities collected at the center are relatively easier to remove than other industrial or manufacturing process.
2. Mass production of symmetrical product was made possible at a much lower cost.
3. It provide dense metal with mechanical soundness.
4. Relatively less temp of molten metal is required for the process saving money and energy.
5. Gates and risers are no longer needed.
6. After the casting process is heavily reduced.

Fig: Separation of impurities toward the center during the whole process.

Types of Centrifugal Casting

The process of using centrifugal force for casting can be divided into three major parts; True centrifugal, semi-centrifugal and centrifuging.

1) True Centrifugal Casting



Centrifugal casting

True centrifugal casting or commonly known as normal centrifugal casting is used to produce a symmetrical hollow structure with round holes. The key feature of this process is to produce a symmetrical hollow structure without using any cores. It is achieved by pure centrifugal force by rotating mold about its vertical or horizontal axis.

The shape of the mold can be either circular, square, rectangular or hexagonal; as long as they are symmetrical about its vertical or horizontal axis of rotation.

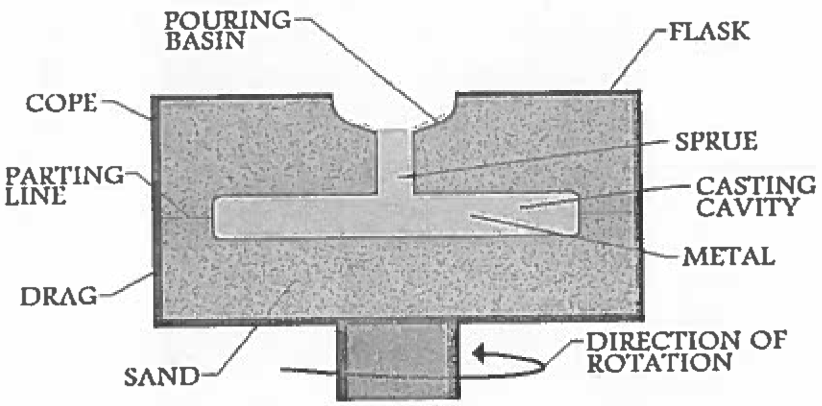
Centrifugal force acting on the molten metal introduced in the system; force it towards the wall of the mold/die. The casting of long parts such as pipes and liners are done along the horizontal axis while for others along the vertical axis.

To avoid molten metal to take the parabolic path along the mold while hardening due to gravity; the mold is subjected to high speed rotation to produce centrifugal force 100 times stronger than of gravity.

The mold is then left to rotate on its axis till it solidifies unidirectionally. This process is generally used for producing large and medium-sized parts such as cylinder liner, hollow pipes, and bushes.

2) Semi-Centrifugal Casting

SEMICENTRIFUGAL CASTING
SOLIDIFICATION OF A WHEEL

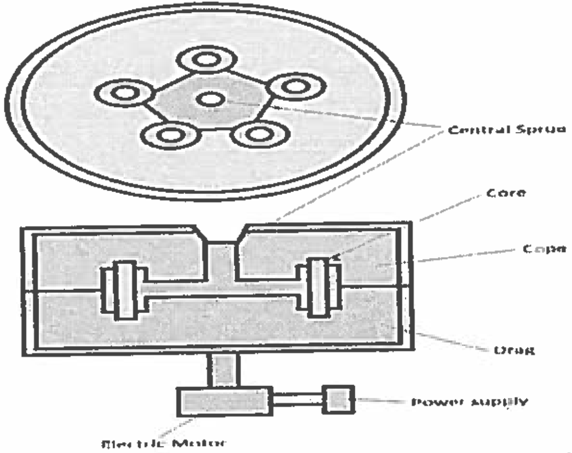


A core is placed inside the mold in a semi centrifugal casting to produce a hollow structure. The process itself includes the rotation of mold along its vertical axis with the core inserted at the center.

The centrifugal force from the rotation is then used to fill the mold perfectly. Here the hot metal is poured along the axis to first fill along the walls with the centrifugal force and towards center due to gravity.

A core is inserted whenever a hollow structure is needed to compose compensating the gravitational forces (dominant) at the center. This process is used to produce large Axis symmetrical products such as flywheel and gear blanks.

3) Centrifuging



This casting process is used to cast metal under high pressure for relatively small mold/die / final product or asymmetrical products in group.

The process is done in group to obtain overall symmetry of the casting. Hot metal is poured into the die along the central axis through central spur using centrifugal force through radial in gates.

It is an easy method for obtaining unidirectional solidification at economical costs.

9(b). Which centrifugal casting method is used for which applications? (6M)

Centrifugal casting provides high material soundness and is the metal casting process of choice for jet engine compressor cases, petrochemical furnace tubes, many military and defense components, and other applications requiring high reliability.

In centrifugal casting molten metal is poured into a spinning die, which can rotate on a vertical axis (vertical centrifugal casting) or horizontal axis (horizontal centrifugal casting) depending on the configuration of the part. Ring and cylinder type shapes are made by vertical centrifugal casting while tubular shapes are made by horizontal centrifugal casting. Either process can be used to produce multiple parts from a single casting.

High centrifugal force applied to the molten metal in the spinning die causes less dense material such as oxides and impurities to “float” to the inner diameter (I.D.) where they concentrate and are removed by machining. Solidification is managed directionally under pressure from the outer diameter (O.D.) to the I.D., avoiding mid-wall shrink. This results in a defect-free structure without cavities or gas pockets.

With the world’s largest and most diverse inventory of centrifugal dies, MetalTek minimizes both upfront tooling costs and product lead times for our customers. We can produce horizontal centrifugal castings with O.D. up to 60" (1,524 mm), length up to 432" (10,973 mm), and weight up to 135,000 lbs. (61,235 kg). Vertical centrifugal castings are available with O.D. up to 180” (4,572 mm) and weight up to 34,000 lbs. (15,422 kg).

10(a). How submerged arc welding process takes place? (7M)

Submerged Arc Welding (SAW) is a joining process that involves the formation of an electric arc between a continuously fed electrode and the workpiece to be welded. A blanket of powdered flux surrounds and covers the arc and, when molten, provides electrical conduction between the metal to be joined and the electrode. It also generates a protective gas shield and a slag, all of which protects the weld zone.

The make-up of the process can be viewed by reference to Figure 1 below

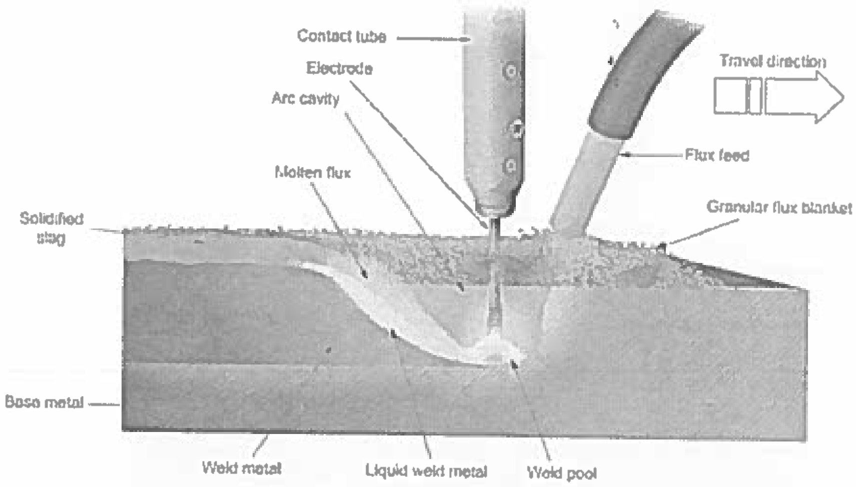


Figure The Submerged Arc Welding Process

As can be seen from Figure , the arc is "submerged" beneath a blanket of flux and is, therefore, not usually visi:ble during the welding operation itself. These facts make the process advantageous from a health and safety viewpoint as there is no arc to promote "arc eye" and very little fume.

There are two welding consumables involved in the process, the electrode and the flux. The electrode can be a solid wire, a cored wire, or a strip. The flux, made from a variety of minerals and compounds, can be rather complex and can be produced in a number of forms.

The general arrangement of the power source and controls, wire feed and flux dispensing are shown in Figure 2.

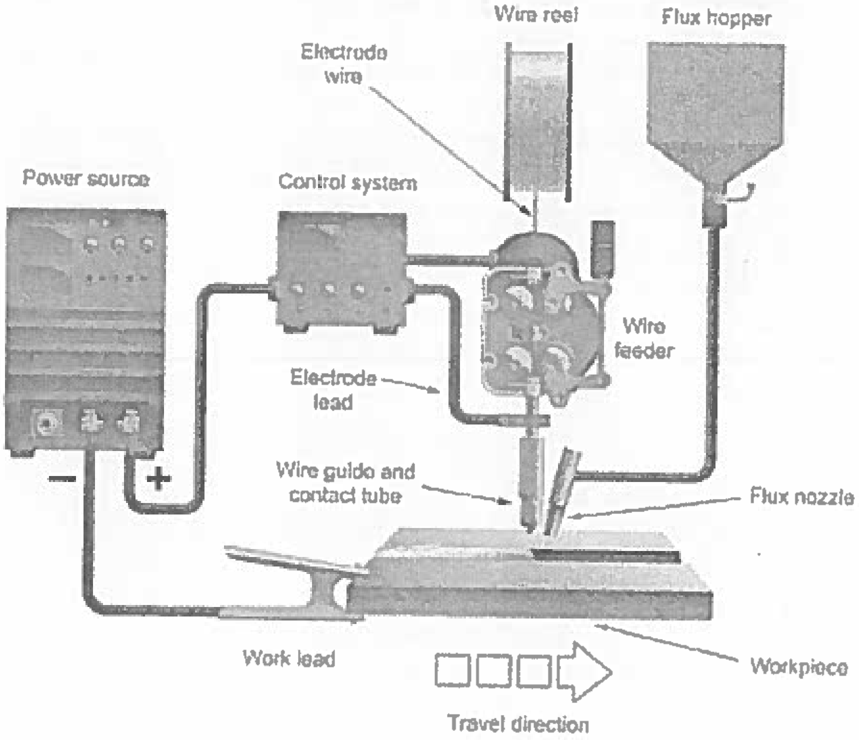


Figure 2. General Arrangement of the Submerged Arc Process

Submerged arc welding is viewed as a high productivity process and is usually automated/mechanized in its form. The simplest application of the process uses a single wire.

Selecting the correct wire diameter for a welded joint depends on many factors and the size of the available power source usually limits the diameter of the wire that can be used. While most power sources for this process are 1,000 amps, smaller power source may be used. A 3/32-in.-dia. wire through to a 5/32-in.-dia. wire will run in the 300 to 900 amps range using direct current and with the electrode positive (DC+)
This welding process is typically suited to the longitudinal and circumferential butt welds required in the manufacture of pressure vessels and for joining plating and stiffeners in shipyards. Welding is positionally restricted and is normally carried out in the flat or horizontal positions because of the highly fluid weld pool, the molten slag, and the need to maintain a flux covering over the arc.
As with all welding processes the selection of the consumables (wire and flux) and other parameters such as amps, volts and travel speed are intended to give a weld deposit that satisfies the objectives of the designer. In the case of this welding process, since the arc is submerged, the welding operator cannot see the molten weld pool and must, therefore, very accurately set the welding parameters and location of the welding nozzle within the joint.

10(b). Describe its applications and advantages? (5M)

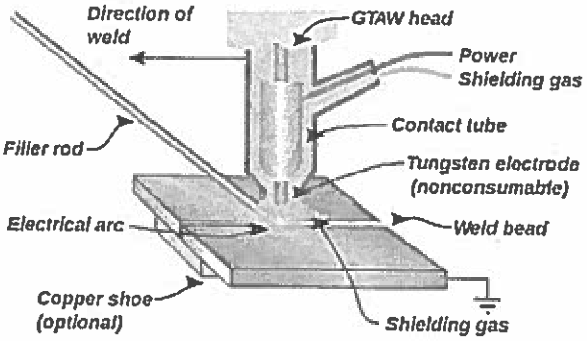
Submerged arc welding has many advantages but there are also restrictions, some of these are listed below
Advantages

- High deposition rates and high arc on times when fully automated.
- Minimal welding fume, no weld spatter and no visible arc
- Unused flux can be recovered
- If metallurgically acceptable, single pass welds can be made in relatively thick plates.

here are many more applications of this welding process, other than its use with a single wire, and the fluxes used can be quite complex in their design and production. These items may be covered in later articles.

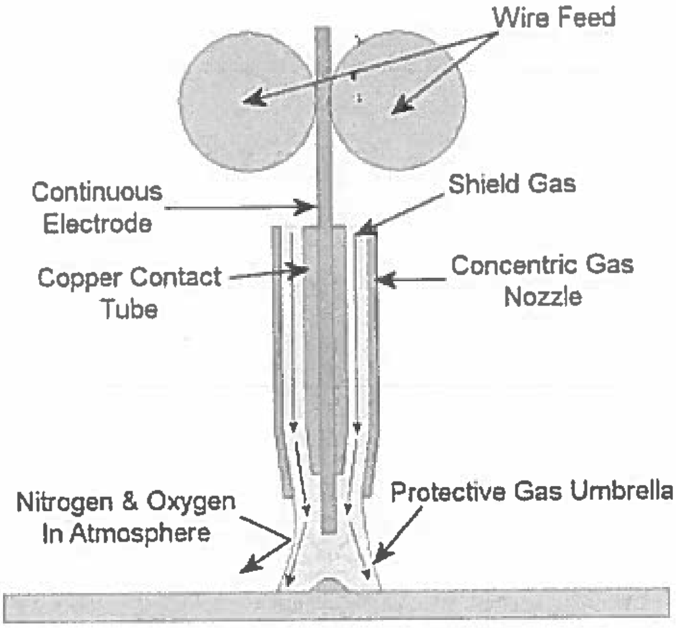
(OR)

11(a). Explain the TIG systems in arc welding. Give its applications of each? (6M)



tungsten Inert Gas (TIG) welding uses the heat generated by an electric arc struck between a non-consumable tungsten electrode and the workpiece to fuse metal in the joint area and produce a molten weld pool. The arc area is shrouded in an inert or reducing gas shield to protect the weld pool and the non-consumable electrode. The process may be operated autogenously, that is, without filler, or filler may be added by feeding a consumable wire or rod into the established weld pool. TIG produces very high quality welds across a wide range of materials with thicknesses up to about 8 or 10mm. It is particularly well suited to sheet material.
The success of this welding process hinges on various factors such as the choice of shielding gas, welding wire, tungsten electrode and the welding technique.

11(b). Explain the MIG systems in arc welding. Give its applications of each? (6M)



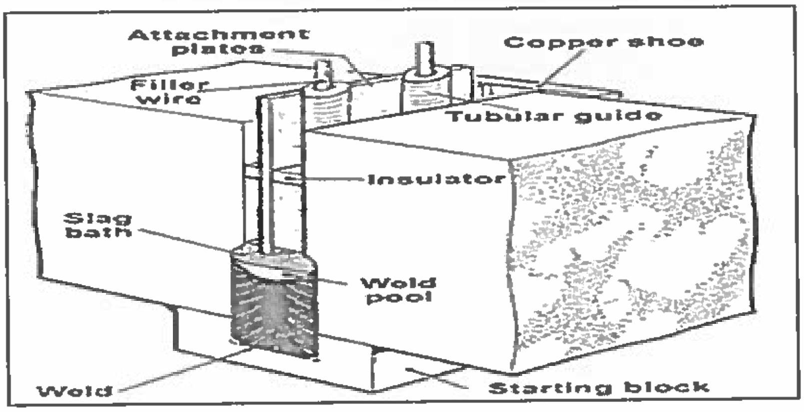
MIG/MAG welding is a versatile technique suitable for both thin sheet and thick section components. An arc is struck between the end of a wire electrode and the workpiece, melting both of them to form a weld pool. The wire serves as both heat source (via the arc at the wire tip) and filler metal for the welding joint. The wire is fed through a copper contact tube (contact tip) which conducts welding current into the wire. The weld pool is protected from the surrounding atmosphere by a shielding gas fed through a nozzle surrounding the wire. Shielding gas selection depends on the material being welded and the application. The wire is fed from a reel by a motor drive, and the welder moves the welding torch along the joint line. Wires may be solid (simple drawn wires), or cored (composites formed from a metal sheath with a powdered flux or metal filling). Consumables are generally competitively priced compared with those for other processes. The process offers high productivity, as the wire is continuously fed.

12(a). Describe the electro slag welding process? (5M)

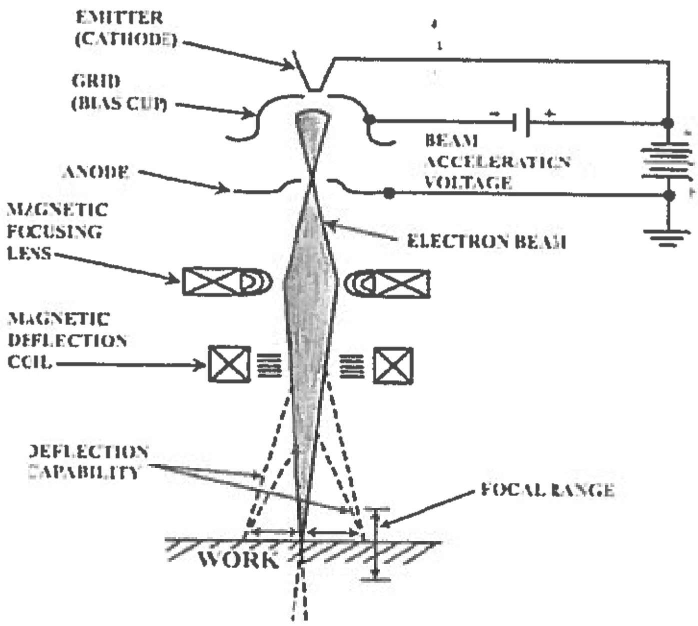
Electroslag Welding is a welding process, in which the heat is generated by an electric current passing between the consumable electrode (filler metal) and the work piece through a molten slag covering the weld surface.

Prior to welding the gap between the two work pieces is filled with a welding flux. Electroslag Welding is initiated by an arc between the electrode and the work piece (or starting plate). Heat, generated by the arc, melts the fluxing powder and forms molten slag. The slag, having low electric conductivity, is maintained in liquid state due to heat produced by the electric current.

The slag reaches a temperature of about 3500°F (1930°C). This temperature is sufficient for melting the consumable electrode and work piece edges. Metal droplets fall to the weld pool and join the work pieces. Electroslag Welding is used mainly for steels.



12(b). Describe the electron beam welding process? (7M)



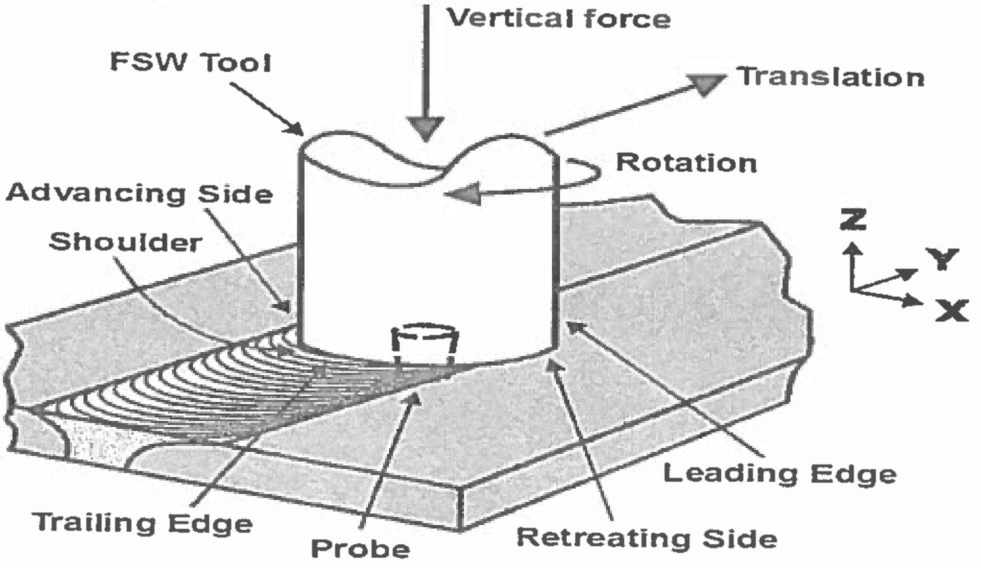
Electron beam (EB) welding is a fusion welding process whereby electrons are generated by an electron gun and accelerated to high speeds using electrical fields. This high speed stream of electrons is tightly focused using magnetic fields and applied to the materials to be joined. The beam of electrons creates kinetic heat as it impacts with the workpieces, causing them to melt and bond together.

Electron beam welding is performed in a vacuum environment as the presence of gas can cause the beam to scatter. Due it being a vacuum process and because of the high voltages used, this welding method is heavily automated and computer controlled. As a result, specialised fixtures and CNC tables are used to move the workpieces inside the welding vacuum chamber.

Recent developments in electron beam welding machine technology have realised a local method of electron beam welding, whereby the electron beam gun is enclosed in a vacuum box on the side of the material to be joined, rather than placing the entire workpiece inside a vacuum chamber.

(OR)

13(a). Explain the method and applications of friction stir welding? (6M)



Friction stir welding (FSW) is a solid-state joining process that uses a non-consumable tool to join two facing workpieces without melting the workpiece material. Heat is generated by friction between the rotating tool and the workpiece material, which leads to a softened region near the FSW tool. While the tool is traversed along the joint line, it mechanically intermixes the two pieces of metal, and forges the hot and softened metal by the mechanical pressure, which is applied by the tool, much like joining clay, or dough It is primarily used on wrought or extruded aluminium and particularly for structures which need very high weld strength. FSW is capable of joining aluminium alloys, copper alloys, titanium alloys, mild steel, stainless steel and magnesium alloys. More recently, it was successfully used in welding of polymers.[3] In addition, joining of dissimilar metals, such as aluminium to magnesium alloys, has been recently achieved by FSW

13(b). Explain any two destructive testing of welds? (6M)

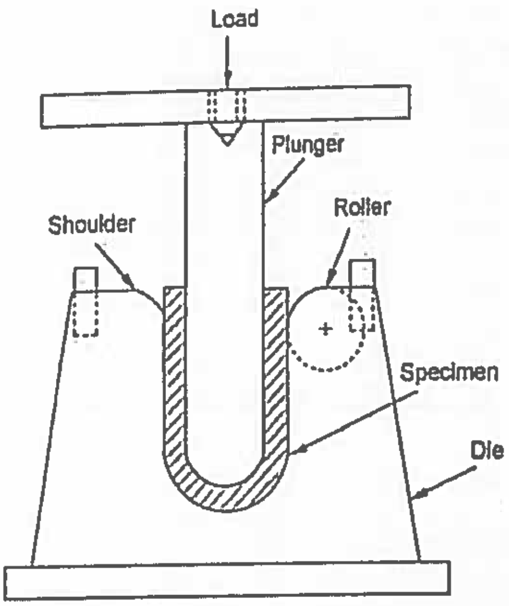
1. Acid Etch Test

This physical weld testing is employed to ascertain the soundness of the weld. The acid attacks the edge of the defects in base metal or weld metal and identifies the weld defects. In the condition of the defect, the boundary becomes accentuated between base and weld metals and can define the defect clearly which is otherwise not visible to the naked eye. This test is performed along the cross-section of the weld joint.

The acid solutions used here are hydrochloric acid, ammonium persulfate, nitric acid, or iodine, and potassium iodide for the etching of carbon and low alloy steel.

2. Guided Bend Test

These guided bend tests are used to determine the quality of the weld metal at the root and face of the welded joint. They also judge the fusion and degree of penetration to the base metal along with the efficiency of the weld. The testing of this type can be done in a jig. The required specimens for testing are machined from the already welded plates, the thickness of these specimens should be within the capacity of our jig for bending. The specimen for testing is placed upon the supports of the die that is the lower part of the jig. The hydraulic jack's plunger forced the specimen into it and assured the shape of the die seen.

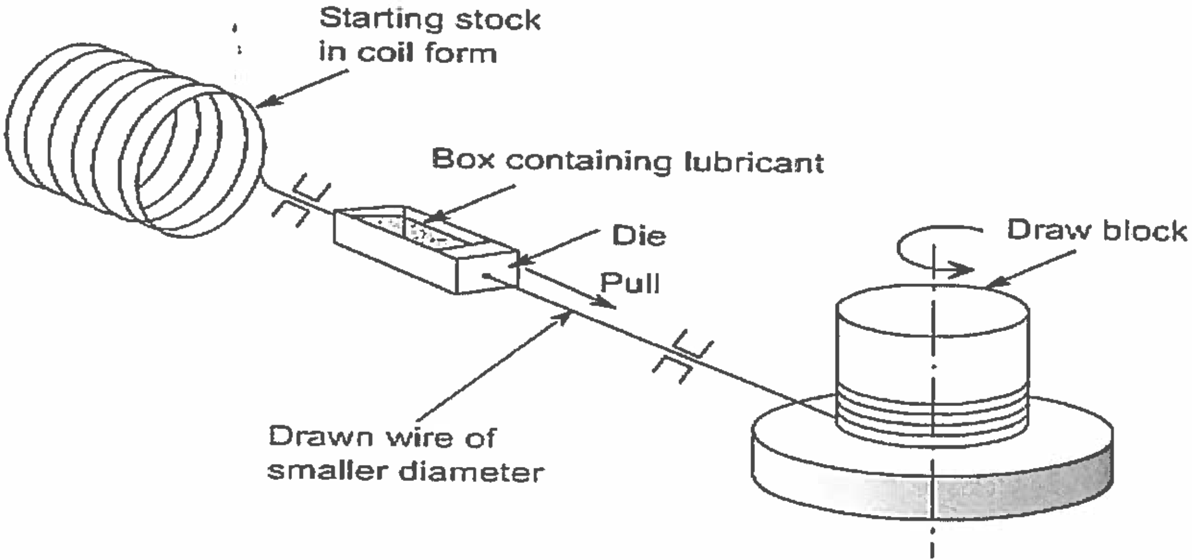


The requirement of this test is fulfilled by bending the specimens at 180 degrees and now accepted as passable. No, any crack more than 3.2mm in any dimension should be visible on the surface. Face bend tests are made in the jig while facing the weld in tension means outside of the bend. Now the root bend test is made in the jig with the face of the weld in tension as on the outside of the bend. The guided bend tests are shown in the figure.

Notes:

- T-Test plate thickness
- A hardened roll may be utilized on shoulders if needed
- Specific dimension for 3/7 of the plate
- Every shown dimension is in inches.

14(a). Describe wire drawing process? (7M)

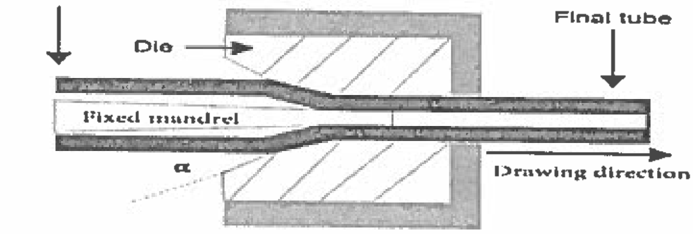


wire drawing, Making of wire, generally from a rod or bar. The wire-drawing process consists of pointing the rod, threading the pointed end through a die, and attaching the end to a drawing block. The block, made to revolve by an electric motor, pulls the lubricated rod through the die, reducing it in diameter and increasing its length. Fine wire is made by a multiple-block machine, because the reduction cannot be performed in a single draft.

14(b). Describe the tube drawing process? (5M)

Tube drawing is a metalworking process used to create a tube with a smaller diameter by pulling, or drawing, a larger diameter tube through a die. There are five methods of tube drawing that are commonly used. These methods are fixed plug drawing, floating plug drawing, tethered plug drawing, rod drawing and tube sinking.

This process is a cold-working process, meaning that the metal tubing is not heated prior to being shaped in the tube drawing process. This gives the finished product added strength because the metal tubing is not affected by thermal expansion during the process. In addition, this process produces tubing with more precise measurements than other methods of production.



(OR)

15(a). Enumerate the typical applications of cold drawing? (4M)

These include pump parts as well as valve stems, linear guide rails or sprockets. Furthermore, they can be gears, keyways, splines and spindles. Turbine parts, curtain wall and facade constructions are also possible. They even include x-ray equipment, louvers, and couplings.

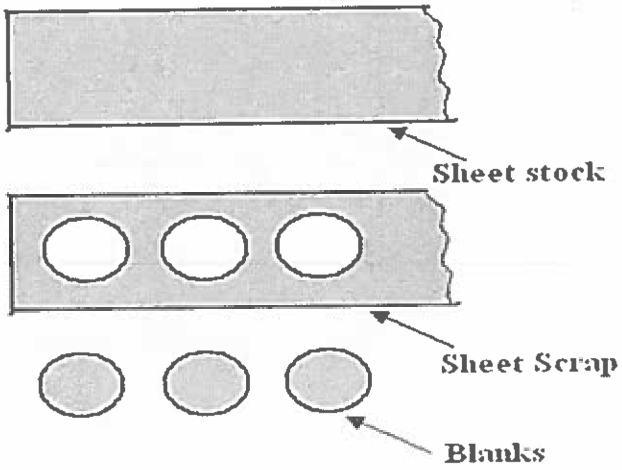
15(b). Explain the blanking and piercing process with a neat sketch? (8M)

Blanking

Punching or blanking is a process in which the punch removes a portion of material from the larger piece or a strip of sheet metal. If the small removed piece is the useful part and the rest is scrap, the operation is called blanking

The piece cut out is called as blank and may be further processed. Blanks are often cut out of a sheet or strip.

Blanking wastes certain amount of material. When designing a sheet metal blanking process the geometry of blanks should be nestled as efficiently as possible to minimize the material waste.



BLANKING

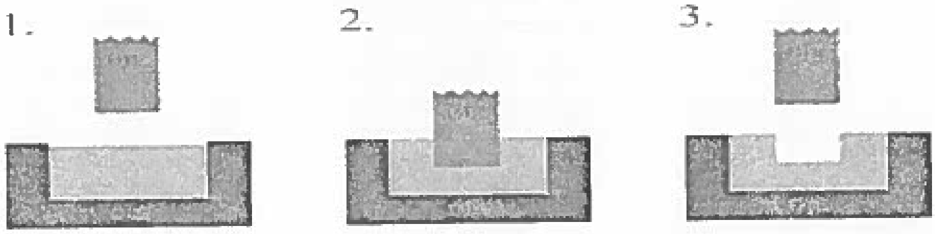
Piercing

It is a process by which a hole is cut (or torn) in metal. It is different from punching in that piercing does not generate a slug. Instead, the metal is pushed back to form a jagged flange on the back side of the hole.

A pierced hole looks somewhat like a bullet hole in a sheet of metal.

Size of the component is generally larger in piercing than blanking.

PIERCING



NSRIT

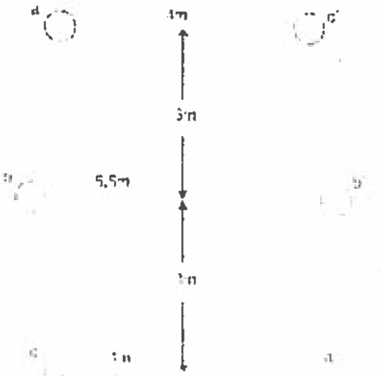
Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	: 2021 - 2022
Course Code	20EE305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	POWER GENERATION AND TRANSMISSION						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Recall Nuclear fission.	20EE305.1	L1
2	Define three part tariff.	20EE305.2	L1
3	Compare GMR and GMD	20EE305.3	L2
4	Classify the transmission lines based on voltage.	20EE305.4	L1
5	List the various methods for improving string efficiency	20EE305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

<table>
<tr><th>No.</th><th>Questions (6 through 15)</th><th>Marks</th><th>Learning Outcome (s)</th><th>DoK</th></tr>
<tr><td>6</td><td>Describe the different blocks in a thermal power plant with neat block diagram. Also mention the importance of boiler accessories economizer and condenser with neat diagrams.</td><td>12M</td><td>20EE305.1</td><td>L2</td></tr>
<tr><td></td><td>OR</td><td></td><td></td><td></td></tr>
<tr><td>7</td><td>Sketch the layout of Hydro power plant and briefly explain the main components and operation of hydro power station.</td><td>12M</td><td>20EE305.1</td><td>L2</td></tr>
<tr><td>8</td><td>A generating station has the following daily load cycle
<table>
<tr><td>Time (Hours)</td><td>0—6</td><td>6—10</td><td>10—12</td><td>12—16</td><td>16—20</td><td>20—24</td></tr>
<tr><td>Load (MW)</td><td>40</td><td>50</td><td>60</td><td>50</td><td>70</td><td>40</td></tr>
</table>
Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor.</td><td>12M</td><td>20EE305.2</td><td>L3</td></tr>
<tr><td></td><td>OR</td><td></td><td></td><td></td></tr>
<tr><td>9</td><td>A generating station has a maximum demand of 20 MW, a load factor of 60%, a plant capacity factor of 48% and a plant use factor of 80% . Find :
i. the daily energy produced
ii. the reserve capacity of the plant
iii. the maximum energy that could be produced daily if the plant was running all the time.
iv the maximum energy that could be produced daily if the plant was running fully loaded and operating as per schedule.</td><td>12M</td><td>20EE305.2</td><td>L3</td></tr>
<tr><td>10</td><td>Derive the calculation of capacitance for 2-wire system in overhead transmission lines.</td><td>12M</td><td>20EE305.3</td><td>L2</td></tr>
<tr><td></td><td>OR</td><td></td><td></td><td></td></tr>
<tr><td>Pg-218 11</td><td>Find the inductance per phase per km of double circuit 3-phase line shown in Figure. The conductors are transposed and are of radius 0·75 cm each. The phase sequence is ABC.
 </td><td>12M</td><td>20EE305.3</td><td>L3</td></tr>
</table>

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12	Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.	12M	20EE305.4	L2
	OR			
13	A 3-phase, 50-Hz overhead transmission line 100 km long has the following constants : Resistance/km/phase = 0.1 Ω Inductive reactance/km/phase = 0·2 Ω Capacitive susceptance/km/phase = $0{\cdot}04 \times 10^{-4}$ siemen Determine (i) the sending end current (ii) sending end voltage (iii) sending end power factor and (iv) transmission efficiency when supplying a balanced load of 10,000 kW at 66 kV, pf. 0·8 lagging. Use nominal T method.	12M	20EE305.4	L3
14	A 3-phase transmission line is being supported by three disc insulators. The potentials across top unit (i.e., near to the tower) and middle unit are 8 kV and 11 kV respectively. Calculate (i) the ratio of capacitance between pin and earth to the self-capacitance of each unit (ii)the line voltage and (iii) string efficiency.	12M	20EE305.5	L2
	OR			
15 (a)	Explain any two types of insulators with neat sketch.	6M	20EE305.5	L2
15 (b)	Derive the expression for the Sag in horizontal plane with equal level supports.	6M	20EE305.5	L2

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SUPPLEMENTARY EXAMINATION KEYS

PART-A SHORT

1. Recall Nuclear fission.

A. The breaking up of nuclei of heavy atoms into two nearly equal parts with release of huge amount of energy is known as nuclear fission.

2. Define three part tariff.

~~A. When the rate of electrical energy is charged on the basis of maximum demand of the Consumer and the units consumed, it is called a two~~

A. When the total charge to be made from the Consumer is split into three part

- fixed charge
- Semi ~~charge~~ fixed. charge
- running charge.

it is known as three part tariff.

3. Compare GMR and GMD

GMR : 1) It is also called Self GMD. (D_s)

2) ~~It is~~ The full form of GMR is Geometrical mean radius

GMD: 1) It is also called as Mutual GMD is the geometrical mean of the distances form one Conductor to the other and. therefore, must be between the largest and Smallest Such distance.

2) In fact, mutual - GMD Simply repr

4. Classify the transmission lines based on voltage.

A.

Classification	Voltage Range (kv)
Extra super voltage Cable.	Beyond 132kV
Extra high tension Cable.	From 33 kV to 66 kv
Super tension Cable	From 22 kv to 33 kv
High tension Cable	From 1 kv to 11 kv
Low tension Cable	upto to 1kv

5. List the Various methods for improving string ~~and~~ Efficiency

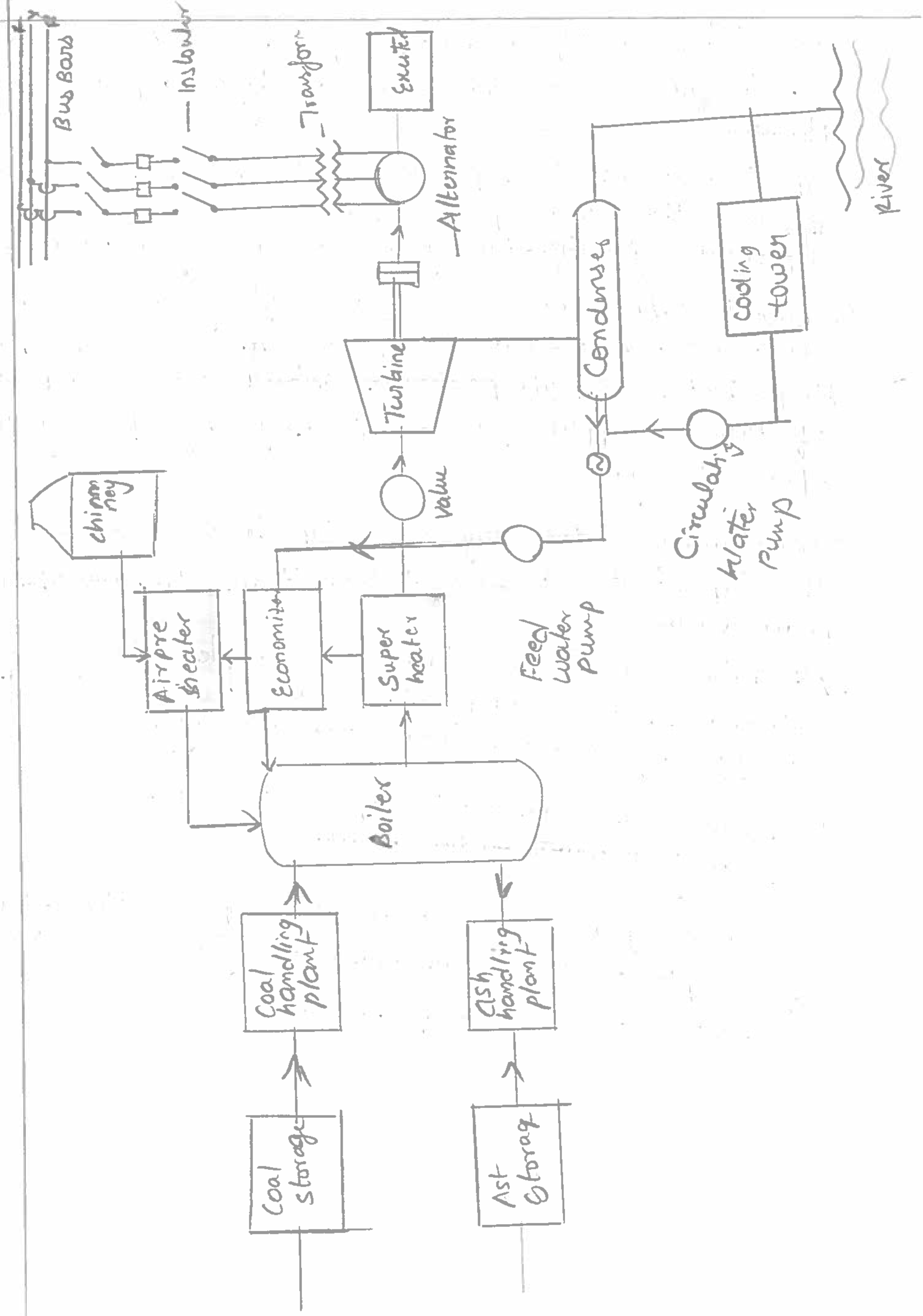
A. 1) By using longer Cross-arms

2) By grading the insulators

3) By using a guard ring

PART-B

6. Describe the different blocks in a thermal power plant with neat block diagram. Also mention the impoltance of boiler accessories economizer and Condenser with neat diagram.


Bus Bars
Insulator
Transfor
Exciter
Alternator
Turbine
Condenser
Cooling tower
River
Circulating Water pump
Feed water pump
Value
chimm ney
Airpre heater
Economiser
Super heater
Boiler
Coal handlig plant
ash handlig plant
Coal storage
Ast Storag

Boiler: The heat of Combustion of Coal in the boiler is utilised to Convert water into steam at high tempeiate and. Pressure.

Super heater: The Steam produced in the boiler is wet and is passed through a Super heater where it is dried and Superheated by the flue gases on their way to chimney.

Economiser: An economiser is essentially a feed water heater and derives heat from the flue gases for this purpose.

Air preheater: An air preheater increases the temperature of the air Supplied for Coal burning by deriving heat from flue gases.

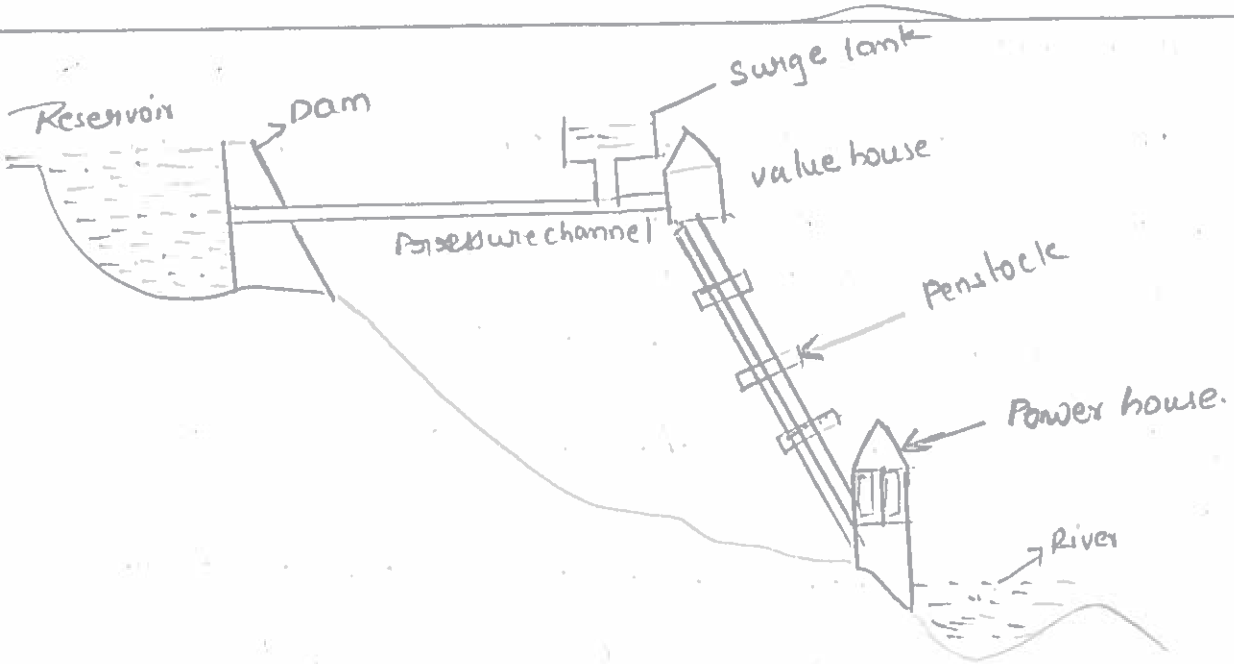
Steam turbine: The dry and Super heater Steam from the Super heater is fed to the Steam turbine through main value.

Alternator: The Steam turbine is Coupled to an alternator. The steo alternator Converts mechanical energy of turbine into electrical energy.

Feed water: The Condensate from the Condenser is used as feed water to the boiler.

Cooling arragngement: In order to improve the Efficiency of the plant, the steam exhausted from the turbine is condensed by means of Condenser.

7. layout of Hydro power plant.



(i) Dam: A dam is a barrier which stores water and creates water head.

(ii) Spillways. There are times when the rivers flow exceeds the storage capacity of the reservoir.

(iii) Head works: The head works consists of the diversion structures at the head of an intake They generally include booms and rack for diverting floating debries and. Sluices for by-passing debris and sediments and. values for ~~diverting for~~ Controlling the flows of water to the turbines.

(iv) Surge tank: Open conduits leading water to the turbine require no Protection. ~~However, w~~ However, when closed conduits are used, protection becomes ~~Acts~~ nec

(v) Penstocks: Penstocks are open or closed conduit which carry water to the turbines.

(vi) water turbine: Water turbines are used to converts the energy of falling water into mechanical energy. and turbine was coupled to an alternator and converts the mechanical energy to electrical energy,

8. Given.data

(i) It is Clear from the load curve that maximum demand on Power Station is 70MW. and occurs during the period 16-20 hours.

∴ Maximum demand = 70MW.

(ii) units generated /day: Area (in kwh) under the load Curve

$= 10^3 [40\times6+50\times4+60\times2+50\times4+70\times4+40\times4]$

$= 10^3 [240+200+120+200+280+160]$ kwh

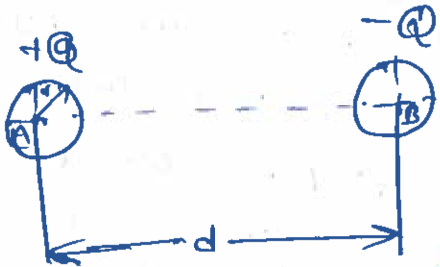
$= 12\times10^5$ kwh.

(iii) Average load $= \dfrac{\text{Units generated /day}}{24\text{hours}} = \dfrac{12\times10^5}{24} = 50,000$ kw

(iv) Load factor: $\dfrac{\text{Average load}}{\text{Max. demand}} = \dfrac{50,000}{70\times10^3} = 0.714 = 71.4\%$

~~9. Given~~ data

10. Capacitance for 2 wire System in overhead transmission.



Consider a single phase overhead transmission line consisting of two Parallel Conductors A and B spaced d metres apart in air. Suppose that radius of each conductor is r meters. let their respective charge be +Q and -Q coulombs

$$V_A = \int_{r}^{\infty} \frac{Q}{2\pi x \varepsilon_0}\,dx + \int_{d}^{\infty} \frac{-Q}{2\pi x \varepsilon_0}\,dx$$

$$= \frac{Q}{2\pi \varepsilon_0}\left[\log_e \frac{\infty}{r} - \log_e \frac{\infty}{d}\right] \text{volts} = \frac{Q}{2\pi \varepsilon_0} \log_e \frac{d}{r} \text{ volts}$$

Similarly, P.d. between Conductor B and neatral 'infinite' plane is

$$V_B = \int_{r}^{\infty} \frac{-Q}{2\pi x \varepsilon_0}\,dx + \int_{d}^{\infty} \frac{Q}{2\pi x \varepsilon_0}\,dx$$

$$= \frac{-Q}{2\pi \varepsilon_0}\left[\log_e \frac{\infty}{r} - \log_e \frac{\infty}{d}\right] = \frac{-Q}{2\pi \varepsilon_0} \log_e \frac{d}{r} \text{ volts}$$

Both these potential are w.r.t the same neutual place

$$V_{AB} = 2V_A = \frac{2Q}{2\pi \varepsilon_0} \log_e \frac{d}{r} \text{ volts}$$

$$C_{AB} = Q/V_{AB} = \frac{Q}{\frac{2Q}{2\pi \varepsilon_0} \log_e \frac{d}{r}} \text{ F/m}$$

$$C_{AB} = \frac{\pi \varepsilon_0}{\log_e \frac{d}{r}} \text{ F/m.} \quad //$$

11. Solution.

Given data.

GMR of Conductor $= 0.75 \times 0.7788 = 0.584$ cm

Distance a to b $= \sqrt{3^2 + (0.75)^2} = 3.1$ m

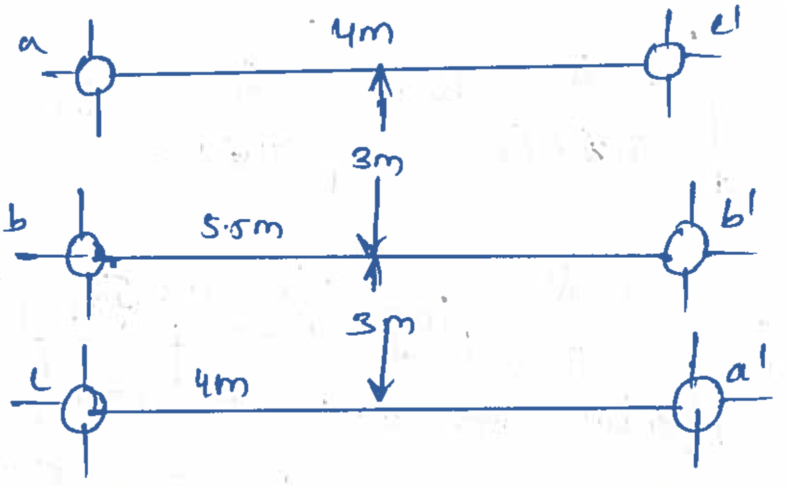
Distance a to b' $= \sqrt{3^2 + (4.75)^2} = 5.62m$

Distance a to a' $= \sqrt{6^2+4^2} = 7.21m$

Equivalent Self GMD of one phase is

$$D_s = \sqrt[3]{D_{S1} \times D_{S2} \times D_{S3}}$$

$$D_{S1} = \sqrt[4]{D_{aa} \times D_{aa'} \times D_{a'a'} \times D_{a'a}}$$



$$= \sqrt[4]{(0.584 \times 10^{-2}) \times (7.21) \times (0.584 \times 10^{-12}) \times (7.21)}$$

$$= 0.205\,m = D_{S3}$$

$$D_{S2} = \sqrt[4]{D_{bb} \times D_{bb'} \times D_{b'b'} \times D_{b'b}}$$

$$= \sqrt[4]{(0.584 \times 10^{-12}) \times (5.5) \times (0.584 \times 10^{-12}) \times 5.5} = 0.18m$$

$$D_s = \sqrt[3]{0.205 \times 0.18 \times 0.205} = 0.195$$

∴ Equivalent mutual G.M.D is

$$D_m = \sqrt[3]{D_{AB} \times D_{BC} \times D_{CA}}$$

$$= D_{AB} = \sqrt[4]{D_{ab} \times D_{ab'} \times D_{a'b} \times D_{a'b'}}$$

$$= \sqrt[4]{3.1 \times 5.62 \times 5.62 \times 3.1}$$

$= 4.17M = D_{BC}$

$D_{CA} = \sqrt[4]{D_{ca} \times D_{ca'} \times D_{c'a} \times D_{c'a'}}$

$= \sqrt[4]{6 \times 4 \times 4 \times 6} = 4.9m$

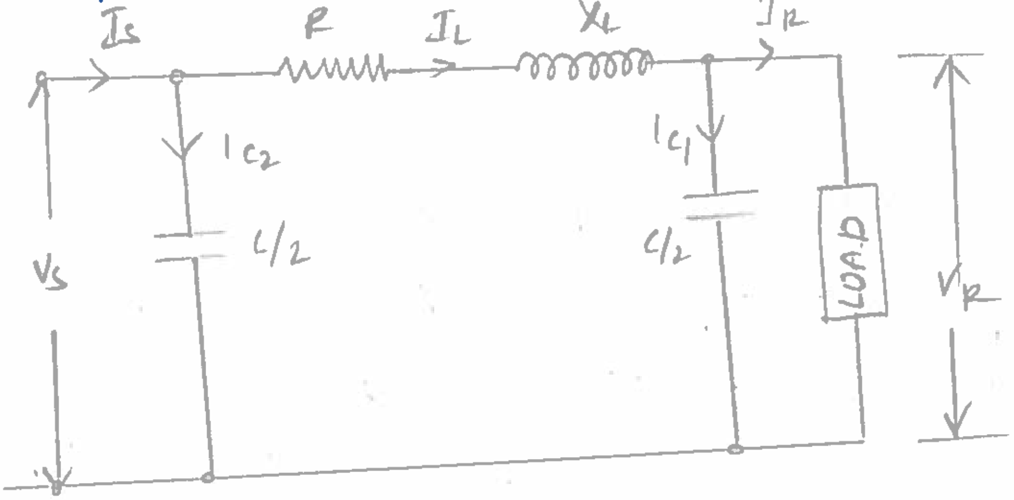
$D_m = \sqrt[3]{4.17 \times 4.17 \times 4.9} = 4.4m$

∴ Inductance /phase /m $= 10^{-7} \times 2 \log_e 4.4/0.195H$

$= 6.23 \times 10^{-7} H = 0.623 \times 10^{-3} mH$

Inductance/phase/km $= 0.623 \times 10^{-3} \times 1000 = 0.623\, mH$.

12. Nominal π method.



In this method, capacitance of each conductor is divided into two halves: one half being lumped at the sending end and other half the receiving end.

Let I_R = load current per phase

R = resistance per phase

X_L = inductive reactance per phase

C = capacitance per phase

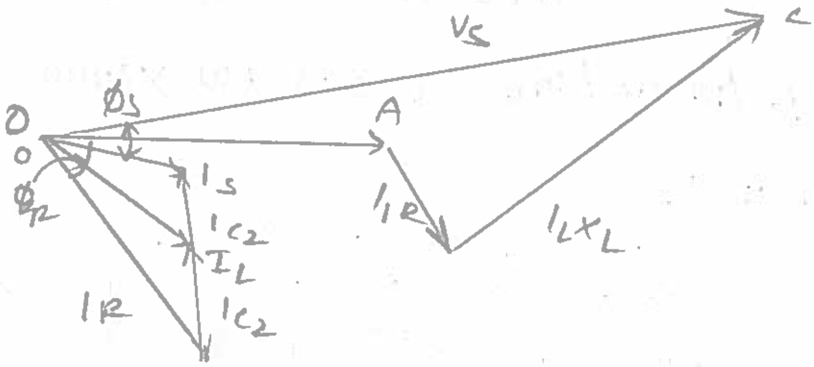
$\cos\phi_R$ = receiving end power factor

V_S = Sending and voltage per phase

The phasor diagram for the Circuit be $\vec{V_R} = V_R + j0$

Load current $\vec{I_R} = I_R(\cos\phi_R - j\sin\phi_R)$

Charging current at load end is $\vec{I_{C_1}} = j\omega(C/2)\vec{V_R} = j\pi f C\vec{V_R}$



Line current $\vec{I_L} = \vec{I_R} + \vec{I_{C_1}}$

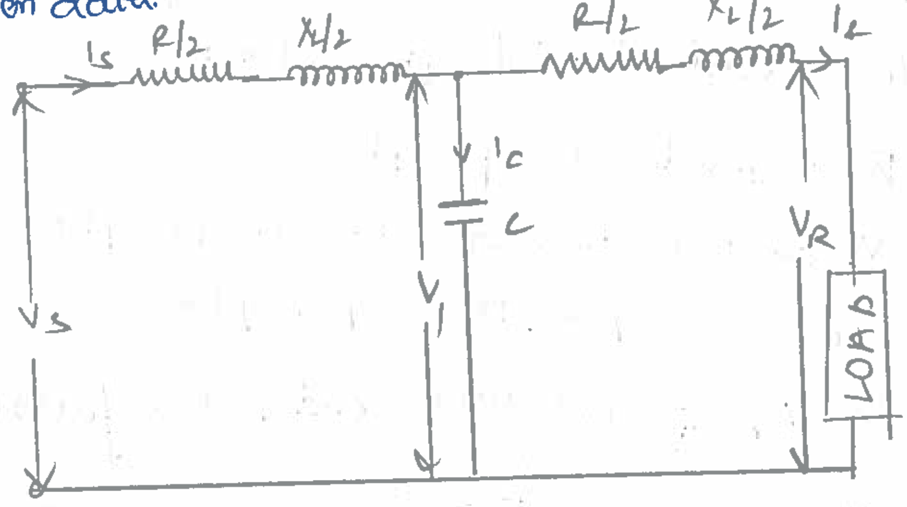
Sending end voltage $\vec{V_S} = \vec{V_R} + \vec{I_L}\vec{Z} = \vec{V_R} + \vec{I_L}(R + jx_L)$

charging Current at the Sending ends

$$\vec{I_{C_2}} = j\omega(C/2)\vec{V_S} = j\pi f C\vec{V_S}$$

∴ sending $\vec{I_S} = \vec{I_L} + \vec{I_{C_2}}$ //

13. Given data.



Total resistance /phase $R = 0.1 \times 100 = 10\Omega$

Total reactance /phase $X_L = 0.2 \times 100 = 20\Omega$

Capacitive susceptance $Y = 0.04 \times 10^{-4} \times 100 = 4 \times 10^{-4} S$

Receiving end Voltage /phase $V_R = 66{,}000/\sqrt{3} = 38105V$

Load current $I_R = \dfrac{10{,}000 \times 10^3}{\sqrt{3} \times 66 \times 10^3 \times 0.8} = 109A$

$\cos\phi_R = 0.8; \ \sin\phi_R = 0.6$

Impedance per phase $\vec{Z} = R + jX_L = 10 + j20$

(i) Taking receiving end voltage as the reference phasor

Receiving end voltage $\vec{V_R} = V_R + j0 = 38{,}105V$

Load Current $\vec{I_R} = I_R(\cos\phi_R - j\sin\phi_R) = 109(0.8 - j0.6$

$= 87.2 - j\,65.4$

Voltage across C, $\vec{V_1} = \vec{V_R} + \vec{I_R}\,\vec{Z}/2 = 38{,}105 + (87.2 - j6$

$(5 + j10)$

$= 38{,}105 + 436 + j872 - j327 + 654$

$= 39{,}195 + j\,545$

Charging Current $\vec{I_C} = jY\vec{V_1} = j4 \times 10^{-4}(39{,}195 + j545$

$= -0.218 + j15.6$

Sending end Current $\vec{I_S} = \vec{I_R} + \vec{I_C} = (87.2 - j65.4)$

$+ (-0.298 + j15.6)$

$= 87.0 - j49.8 = 100\angle -29°47'A$

∴ Sending end Current $= 100A$

(ii) Sending end voltage, $\vec{V_s} = \vec{V_1} + \vec{I_s} \vec{Z}/2 = (39.195 + j545)$

$+ (87.0 - j49.8)(5 + j10)$

$= 39,195 + j545 + 434.9 + j870 - j249 +$

$+498$

$= 40128 + j1170 = 40145 \angle 1°40' V$

∴ Line value of sending end voltage

$= 40145 \times \sqrt{3} = 69533$ kw

(iii) Referring to phasor diagram

θ_1 = angle between $\vec{V_R}$ and $\vec{V_S} = 1°40'$

θ_2 = angle " $\vec{V_R}$ and $\vec{I_s} = 29°47'$

ϕ_s = angle " $\vec{V_s}$ and $\vec{I_s}$

$= \theta_1 + \theta_2 = 1°40' + 29°47' = 31°27$

∴ sending end power factor, $\cos\phi_s = \cos 31°27'$
$= 0.853$ lag

(iv) Sending end power $= 3 V_s I_s \cos\phi_s = 3 \times 40,145 \times 100 \times 0.853$
$= 10273105$ w $= 10273.105$ kw

Power delivered $= 10,000$ kw

∴ Transmission Effiency $= \frac{10,000}{10273,105} \times 100 = 97.34\%$

14. Given data

3phase.

$V_1 = 8 kV = 8 \times 10^3$

$V_2 = 11 kV = 11 \times 10^3$

no of discs = 3 dics.

formula

$V_2 = (1 + 1/m) v_1 =$

$[\therefore \frac{1}{m} = k]$

$V_2 = (1+k)\, V_1 = 11\times10^3 = (1+k)\, 8\times10^3$

$$= \frac{11\times10^3}{8\times10^3} = 1+k$$

$1.375 = 1+k$

$0.375 = 1+k$

$0.375 = k$

$\boxed{K = 0.375}$

$V_3 = \left(1+3/m + 1/m^2\right) V_1 = 1+3(0.375)+(0.375)^2$

$= 2.2656 V_1 = 2.2656 \times 8000 = 18120$

$V_E = V_1+V_2+V_3 = 8000+11000+18120$

$= 37120V$

(ii) Line voltage

—Apply KCL law to junction

$V_1+V_2+V_3 = 8+11+18.12$

$= 37.12$

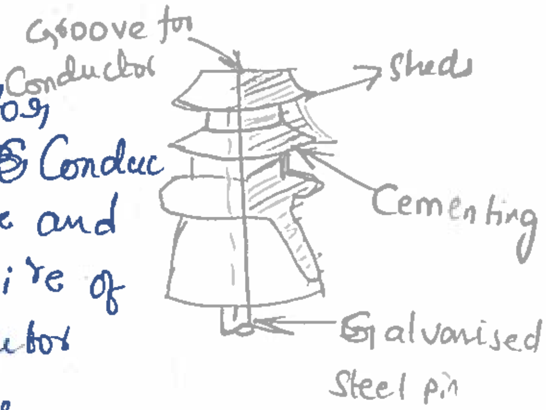
$37.12\times\sqrt{3} = 64.29$ kV.

(iii) String efficiency $= \frac{V}{V_3+n}\times100 = \frac{37120}{18120\times3}$

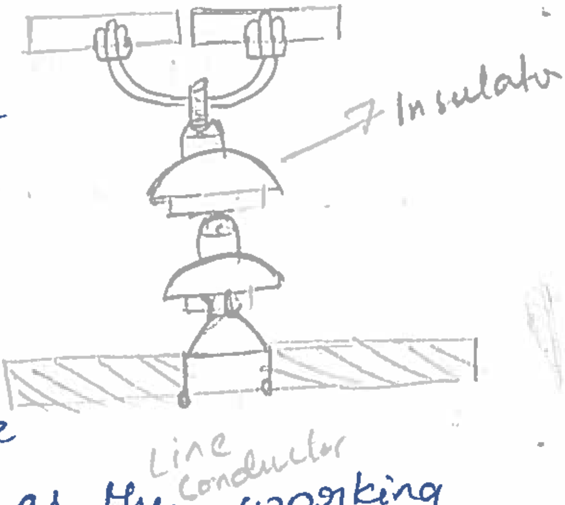
$= 0.6828 \times 100$

$= 68.28\%$

15 (a) Pin type Insulators: The part section of a pin type insulator, the Pin type insulator. is secured Pole. The is a groove on the uper end of the insulator for housing the Conductor. The ~~B~~ Conduc-tor passes through this groove and is bound by the annetleeb wire of the Same material as the Conductor



Pin type insulators are used for transmission and distribution of electric power at voltage upto 33kv



(b)
Suspension type insulators.

The cost of Pin type insulator increases rapidly as the working voltage is increased. Therefore this type of insulator is not economical beyonds 33 kv.

The Consists of a number of porcelain dics Connected in series by metal links in the form of String. The Conductor is Suspended at the bottom end of this String while . the other ~~ed~~end of the String ~~B~~ is secured to the cross-arm of tower.

Advantage

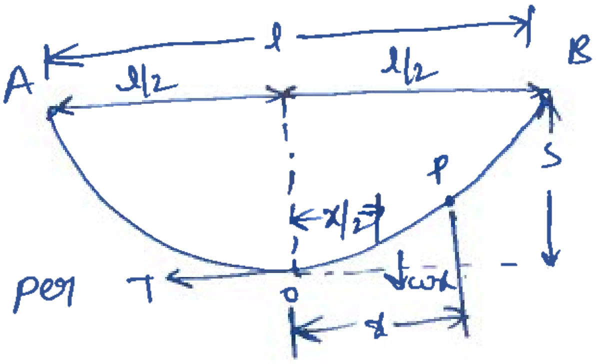
(i) Suspension type insulation are cheaper.

(ii) If any one dics is damaged the whole String does not become useless because

the stisfactory to Supply the greater ~~demand~~ . flexibility to the line

15(b) when Supports are at Equal levels

Consider a Conductor between two equilevel Supports A and B with 'o'



let l = length of the Span

w = weight of the Conductor per meter height

T = Tension of the Conductor

O = lowest point on the Conductor.

S = sag.

Consider a point P on the Conductor at a distance x meter from O

The two force acting on a portion OP

1. Tension 'T' acting at 'o' point

2. weight of the Conductor OP acting down words the at x/2 meter.

So Equating the moment of the above two force at P

$$Ty = wx \cdot \frac{x}{2}$$

$$\Rightarrow \frac{wx^2}{2T} \Rightarrow Y = \frac{wx^2}{2T}$$

$$\boxed{\therefore x = l/2}$$

$$\boxed{Y = S}$$

$$Y = \frac{w(l/2)^2}{2T}$$

$$\boxed{S = \frac{wl^2}{8T}}$$

Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2021 - 2022
Course Code	20EC305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Digital System Design**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	(11x1y)8 =(12c9)16 find x and y values	20EC305.1	L1
2	Implement F= AB+ B'C using logic gates.	20EC305.2	L1
3	Differentiate between RAM and ROM.	20EC305.3	L1
4	State the difference between Latch and Flip-Flop.	20EC305.4	L1
5	What is the basic use of EDA tools?	20EC305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Convert The Following. i. (A6.CE)16 = ()10 ii. (1266.756)8 = ()10 iii. (10100011.11001)2 = ()10 iv (179.897) 10= ()16	6M	20EC305.1	L2
6 (b)	Represent AND, OR, NOT, NAND gate using NOR gate	6M	20EC305.1	L2
	OR			
7 (a)	Determine the single error correct code for the information code 10111 for odd parity	6M	20EC305.1	L2
7 (b)	(i) Convert the following binary 1010011 into gray code (ii)Convert the following gray code 101011 into its equivalent binary	6M	20EC305.1	L2
8 (a)	Demonstrate by means of truth tables the Boolean Associative law and distributive law	6M	20EC305.2	L2
8 (b)	Simplify the following Boolean expression to a minimum number of literals. i. F=(BC1+A1D)(AB1+CD1) ii. F=WYZ+XY+XZ1+YZ.	6M	20EC305.2	L2
	OR			
9 (a)	Obtain the simplified expression in sum of products using K-map method F(A , B , C , D) = Σ(0, 1, 4, 5, 7, 8, 15)	6M	20EC305.2	L2
9 (b)	Obtain the simplified expression in POS (product of sums) of F(w, x,y,z)= Σ (1,2,4,7,12,14,15) using K-maps	6M	20EC305.2	L2
10 (a)	State the applications of demultiplexer and implement 1 to 4 De-Mux.	6M	20EC305.3	L3
10 (b)	Construct 4*16 decoder using two 3*8 decoders	6M	20EC305.3	L3
	OR			
11 (a)	Build a 2 Bit Magnitude Comparator using gates	6M	20EC305.3	L3
11 (b)	Develop the following Boolean functions using PLA with 3 AND gates. F1 (ABC) = Σ (3,5,7), F2 = Σ(4,5,7).	6M	20EC305.3	L3
12 (a)	Mention the drawback of JK flip- flop and explain in what way the drawback is eliminated in Master -Slave JK flip- flop.	6M	20EC305.4	L2
12 (b)	Draw the logic diagram of a SR flip flop using NAND gates and illustrate its operation using truth table and find the expression.	6M	20EC305.4	L2
	OR			

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13 (a)	Show the circuit diagram of 4-bit Johnson counter using D-flip flop and explain its operation with the help of bit pattern	6M	20EC305.4	L2
13 (b)	Draw the block diagram of asynchronous sequential circuit	6M	20EC305.4	L2
14 (a)	Contrast the program structure of VHDL and Explain the significance of entity and architecture	6M	20EC305.5	L4
14 (b)	Inspect a VHDL code for 4:1 MUX using Case Statement	6M	20EC305.5	L4
	OR			
15 (a)	Analyze the dataflow design style of VHDL programming with suitable example	6M	20EC305.5	L4
15 (b)	List out the detail about packages and libraries used in VHDL	6M	20EC305.5	L4

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Scheme & Key For DSD (20EC305)

SEMESTER END SUPPLEMENTARY EXAMINATION, APRIL/MAY [illegible]

Department of ECE

Subject:- Digital System Design　　Course Code: 20EC305
AY:- 2021-22　　Sem - III

Part (A) (Short Answer Questions $5 \times 2 = 10$ M)

1. $(11X1Y)_8 = (12C9)_{16}$ find x & y values — (2M)

Sol:- Converting Hexadecimal to binary

$(12C9)_{16} = 0001001011001001_{(2)}$ — 1M

Now Convert Binary to Octal

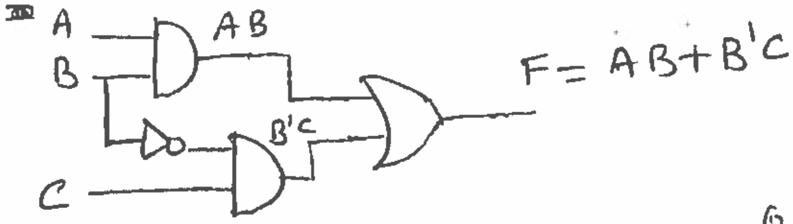
$$\underbrace{00\,0}_{0}\underbrace{001}_{1}\underbrace{001}_{1}\underbrace{011}_{3}\underbrace{001}_{1}\underbrace{001}_{1}$$ — 1M

$\therefore (11X1Y)_8 = (11311)_8$

$\therefore X = 3$ & $Y = 1$

2. Implement $F = AB + B'C$ using logic gates — (2M)

Sol:- Given $F = AB + B'C$



3. Differentiate between RAM and ROM — (2M)

Sol:-

RAM	ROM
1) RAM is a Volatile memory	1) ROM is a non-volatile memory
2) It stores data temporary	2) It stores data Permanently
3) It is high Speed memory	3) It is slower than the RAM
4) It is Costlier	4) It is Cheaper.

4. State the difference between Latch and Flip Flops:(2M).

Sol:-

Latch	Flip Flop
1) It is used to store single bit of data	1) It is also used to store single bit of data
2) Latch is having a Enable	2) Flip Flop is having a Clock
3) It is a level-triggered type	3) It is a edge-triggered type
4) It is Faster	4) It is slower.

5) What is the basic use of EDA tools? (2M)

Sol:- EAD expands to Electronic Design Automation and these tools are used for synthesis, Implementation and simulation of Electronic circuits on the Software itself.

Part B (Long Answer Questions 5X12 = 60 M)

6(a) Convert the following —— (6M)

(i) $(A6.CE)_{16} = (\quad)_{10}$ —— (1.5 M)

Sol:- $A6.CE_{16} = A\times16^1 + 6\times16^0 + C\times16^{-1} + E\times16^{-2}$

$= 10\times16^1 + 6\times16^0 + C(12)\times16^{-1} + (14)E\times16^{-2}$

$= 160 + 6 + 0.75 + 0.05468$

$= 166.80468_{(10)}$

$\therefore (A6.CE)_{16} = (166.80468)_{10}$

(2)
(2)

(ii) $(1266.756)_8 = (\quad)_{10}$ —— (1.5 M)

Sol:- $1266.756_8 = 1\times8^3 + 2\times8^2 + 6\times8^1 + 6\times8^0 + 7\times8^{-1} + 5\times8^{-2} + 6\times8^{-3}$

$= 512 + 128 + 48 + 6 + 0.875 + 0.07812 + 0.0117$

$= (694.9648)_{10}$

$\therefore (1266.756)_8 = (694.9648)_{10}$

(iii) $(10100011.11001)_2 = (\quad)_{10}$ — (1.5 M)

Sol:- $10100011.11001_2 = 1\times2^7 + 0\times2^6 + 1\times2^5 + 0\times2^4 + 0\times2^3 + 0\times2^2 + 1\times2^1 + 1\times2^0 + 1\times2^{-1} + 1\times2^{-2} + 0\times2^{-3} + 0\times2^{-4} + 1\times2^{-5}$

$= 128 + 0 + 32 + 0 + 0 + 0 + 2 + 1 + 0.5 + 0.25 + 0 + 0 + 0.0312$

$= (163.7812)_{10}$

$\therefore (10100011.11001)_2 = (163.7812)_{10}$

(iv) $(179.897)_{10} = (\quad)_{16}$ — (1.5 M)

Sol:- 16 | 176
| 11 − 0

→ Successive Division method

$\therefore 179_{10} = B0_{16}$

$0.897 \times 16 = 14.352 - 14 = 0.352$
$0.352 \times 16 = 5.632 - 5 = 0.632$
$0.632 \times 16 = 10.112 - 10 = 0.112$
$0.112 \times 16 = 1.792 - 1 = 0.792$
$0.792 \times 16 = 12.672 - 12 = 0.672$

→ Successive multiplication method

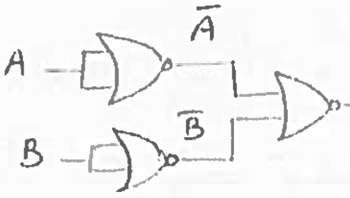
$\therefore 0.897_{10} = E5A1C_{16}$

$\therefore 179.897_{10} = B0.E5A1C_{16}$

6 (b) Represent AND, OR, NOT, NAND gate using NOR gate. (6M)

Sol:- AND Gate

$Y = A \cdot B$

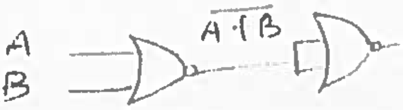


$$Y = \overline{\bar{A} + \bar{B}}$$
$$= \bar{\bar{A}} \cdot \bar{\bar{B}}$$
$$= AB$$

(1.5M)

OR Gate

$Y = A + B$



$$Y = \overline{\overline{A+B}}$$
$$= A + B$$

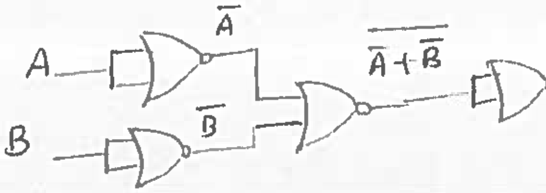
(1.5M)

NOT Gate

$$\overline{X + X} = \bar{X}$$

NAND Gate

$Y = \overline{AB}$



$$Y = \overline{\overline{\bar{A} + \bar{B}}}$$
$$= \bar{A} + \bar{B}$$
$$= \overline{AB}$$

(1.5M)

7(a) Determine the single error correct code for the information code 10111 for odd parity (6M)

Sol:- Given Information Code = 10111

No of Information bits $x = 5$

No of parity bits $p = ?$

$$2^p \geq x + p + 1$$

Let $p=0$, $2^0 \geq 5+0+1$

$\boxed{1 \geq 6}$ Not satisfied

Let $p=1$, $2^1 \geq 5+1+1$

$\boxed{2 \geq 7}$ Not satisfied

Let $p=2$, $2^2 \geq 5+2+1$

$\boxed{4 \geq 8}$

Not satisfied

Let $p=3$

$2^3 \geq 5+3+1$

$\boxed{8 \geq 9}$ Not satisfied

Let $p=4$

$2^4 \geq 5+4+1$

$\boxed{16 \geq 10}$ Satisfied

$\therefore$ No of parity bits $= 4$

∴ Hamming Code length $= x + P$
$= 5 + 4$
$= 9$ bits

Parity bits located in power of 2 i.e $2^0, 2^1, 2^2, 2^3 \ldots$

D_9	P_8	D_7	D_6	D_5	P_4	D_3	P_2	P_1
1	?	0	1	1	?	1	?	?
1001	1000	0111	0110	0101	0100	0011	0010	0001

$P_1 \rightarrow$ Check for odd parity

$P_1 \rightarrow D_3, D_5, D_7, D_9$
$P_1 \rightarrow 1 \quad 1 \quad 0 \quad 1 \rightarrow$ Odd No of 1's

$\therefore P_1 = 0$

$P_2 \rightarrow$ Check for odd parity

$P_2 \rightarrow D_3, D_6, D_7$
$P_2 \rightarrow 1 \quad 1 \quad 0 \rightarrow$ Even No of 1's

$\therefore P_2 = 1$

$P_4 \rightarrow$ Check for odd parity

$P_4 \rightarrow D_5, D_6, D_7 \rightarrow$ Even No of 1's
$P_4 \rightarrow 1 \quad 1 \quad 0$

$\therefore P_4 = 1$

$P_8 \rightarrow$ Check for Odd parity

$P_8 \rightarrow D_9$
$P_8 \rightarrow 1 \rightarrow$ odd No of 1's

$\therefore P_8 = 0$

∴ Hamming Code = 100111110

7(b) (i) Convert the following binary 1010011 into gray Code —(3M)

Sol:- Given binary is 1010011

$$\begin{array}{ccccccc} B_1 & B_2 & B_3 & B_4 & B_5 & B_6 & B_7 \\ 1 & 0 & 1 & 0 & 0 & 1 & 1 \end{array}$$

Binary code is 1010011

$G_1 = B_1$

$G_1 = 1$

$G_2 = B_1 \oplus B_2$

$= 1 \oplus 0$

$= 1$

$G_3 = B_2 \oplus B_3$

$= 0 \oplus 1$

$= 1$

$G_4 = B_3 \oplus B_4$

$= 1 \oplus 0$

$= 1$

$G_5 = B_4 \oplus B_5$

$= 0 \oplus 0$

$= 0$

$G_6 = B_5 \oplus B_6$

$= 0 \oplus 1$

$= 1$

$G_7 = B_6 \oplus B_7$

$= 1 \oplus 1$

$= 1$

∴ The Gray Code is 1111011

7(b) (ii) Convert the following Gray Code 101011 into its equivalent binary (3M)

$$\begin{array}{cccccc} G_1 & G_2 & G_3 & G_4 & G_5 & G_6 \\ 1 & 0 & 1 & 0 & 1 & 1 \end{array}$$

Sol: Given Gray Code is 101011

$B_1 = G_1$

$B_1 = 1$

$B_2 = B_1 \oplus G_2$

$= 1 \oplus 0$

$= 1$

$B_3 = B_2 \oplus G_3$

$= 1 \oplus 1$

$= 0$

$B_4 = B_3 \oplus G_4$

$= 0 \oplus 0$

$= 0$

$B_5 = B_4 \oplus G_5$

$= 0 \oplus 1$

$= 1$

$B_6 = B_5 \oplus G_6$

$= 1 \oplus 1$

$= 0$

∴ The Binary Code is 110010

(i)

8(A) Demonstrate by means of truth tables the Boolean Associative law and Distributive law ---(6M)

Sol:- Associative Law ---(1.5 M)

Law 1:- $(A+B)+C = A+(B+C)$

A	B	C	A+B	(A+B)+C	B+C	A+(B+C)
0	0	0	0	0	0	0
0	0	1	0	1	1	1
0	1	0	1	1	1	1
0	1	1	1	1	1	1
1	0	0	1	1	0	1
1	0	1	1	1	1	1
1	1	0	1	1	1	1
1	1	1	1	1	1	1

---(1.5M)

Law 2:- $(A.B).C = A.(B.C)$

A	B	C	A.B	(A.B).C	(B.C)	A.(B.C)
0	0	0	0	0	0	0
0	0	1	0	0	0	0
0	1	0	0	0	0	0
0	1	1	0	0	1	0
1	0	0	0	0	0	0
1	0	1	0	0	0	0
1	1	0	1	0	0	0
1	1	1	1	1	1	1

(1.5 M

Distributive Law:-

Law 1:- $A+BC=(A+B)(A+C)$

A	B	C	BC	A+BC	A+B	A+C	(A+B)(A+C)
0	0	0	0	0	0	0	0
0	0	1	0	0	0	1	0
0	1	0	0	0	1	0	0
0	1	1	1	1	1	1	1
1	0	0	0	1	1	1	1
1	0	1	0	1	1	1	1
1	1	0	0	1	1	1	1
1	1	1	1	1	1	1	1

Taw2: $A(B+C) = AB + AC$ (1.5M).

A	B	C	B+C	A(B+C)	AB	AC	AB+AC
0	0	0	0	0	0	0	0
0	0	1	1	0	0	0	0
0	1	0	1	0	0	0	0
0	1	1	1	0	0	0	0
1	0	0	0	0	0	0	0
1	0	1	1	1	0	1	1
1	1	0	1	1	1	0	1
1	1	1	1	1	1	1	1

8(b) Simplify the following Boolean expression to a minimum number of literals. - (6M)

(i) $F = (BC' + A'D)(AB' + CD')$

(ii) $F = WYZ + XY + XZ' + YZ$

Sol: (i) $F = (B\bar{C} + \bar{A}D)(A\bar{B} + C\bar{D})$ —(3M)

$= AB\bar{B}C + BC\bar{C}\bar{D} + A\bar{A}\bar{B}D + \bar{A}D\bar{D}C$

$= 0 + 0 + 0 + 0$ [∵ $B\bar{B} = 0$ &, $C\bar{C} = 0$ &, $A\bar{A} = 0$ & $D\bar{D} = 0$]

$= 0$

(ii) $F = WYZ + XY + XZ' + YZ$ (3. M)

$= Y(WZ + X) + XZ' + YZ$ [∵ $A + BC = (A+B)(A+C)$]

$= Y[(X+W)(X+Z)] + XZ' + YZ$

$= Y[(\cancel{X}.X + XZ + XW + WZ)] + XZ' + YZ$

$= Y[(X + XZ + XW + WZ)] + XZ' + YZ$

$= XY + XYZ + XYW + YWZ + XZ' + YZ$

$= XY(1+Z) + XYW + YWZ + XZ' + YZ$ [∵ $1+A=1$]

$= XY + XYW + YWZ + XZ' + YZ$ [∵ $1+A = 1$]

$= XY(1+W) + YWZ + XZ' + YZ$

$= XY + YWZ + XZ' + YZ$

$= XY + YZ + XZ' + YZ$

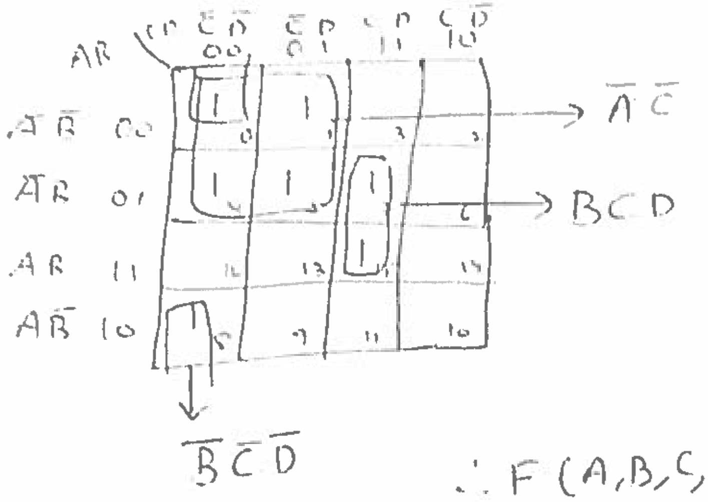
$= XY + YZ(1+Z) + XZ'$ $(\because 1+A=1)$

$= XY + YZ + XZ'$

$= Y(X+Z) + XZ'$

9(a) obtain the simplified expression in SOP using K-map
$F(A,B,C,D) = \Sigma(0,1,4,5,7,8,15)$ — (6M)

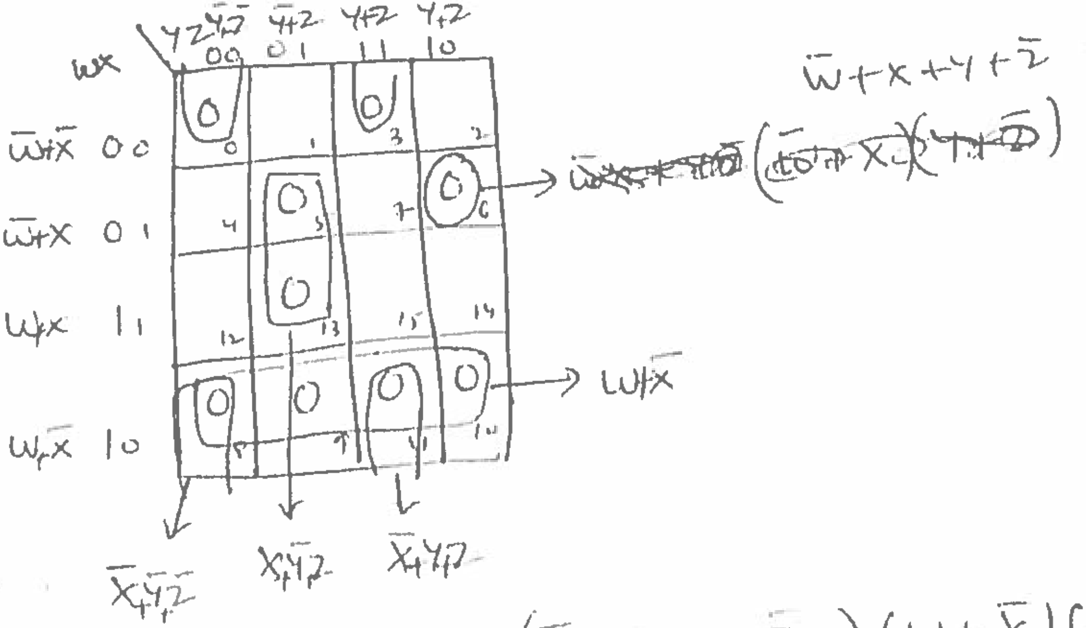
Sol:- $F(A,B,C,D) = \Sigma(0,1,4,5,7,8,15)$ — ~~(1)~~



$\therefore F(A,B,C,D) = \bar{A}\bar{C} + \bar{B}\bar{C}\bar{D} + BCD$

9(b) Obtain the simplified expression in POS of
$F(W,X,Y,Z) = \Sigma(1,2,4,7,12,14,15)$ using K Map
—6M

Sol:- $F(W,X,Y,Z) = \Sigma(1,2,4,7,12,14,15)$



$\therefore F(WXYZ) = (\bar{W}+X+Y+\bar{Z})(W+\bar{X})(\bar{X}+Y+Z)$
$(X+\bar{Y}+Z)(\bar{X}+\bar{Y}+\bar{Z})$

10 (a) State the application of demultiplexer and implement 1 to 4 De-mux — GM

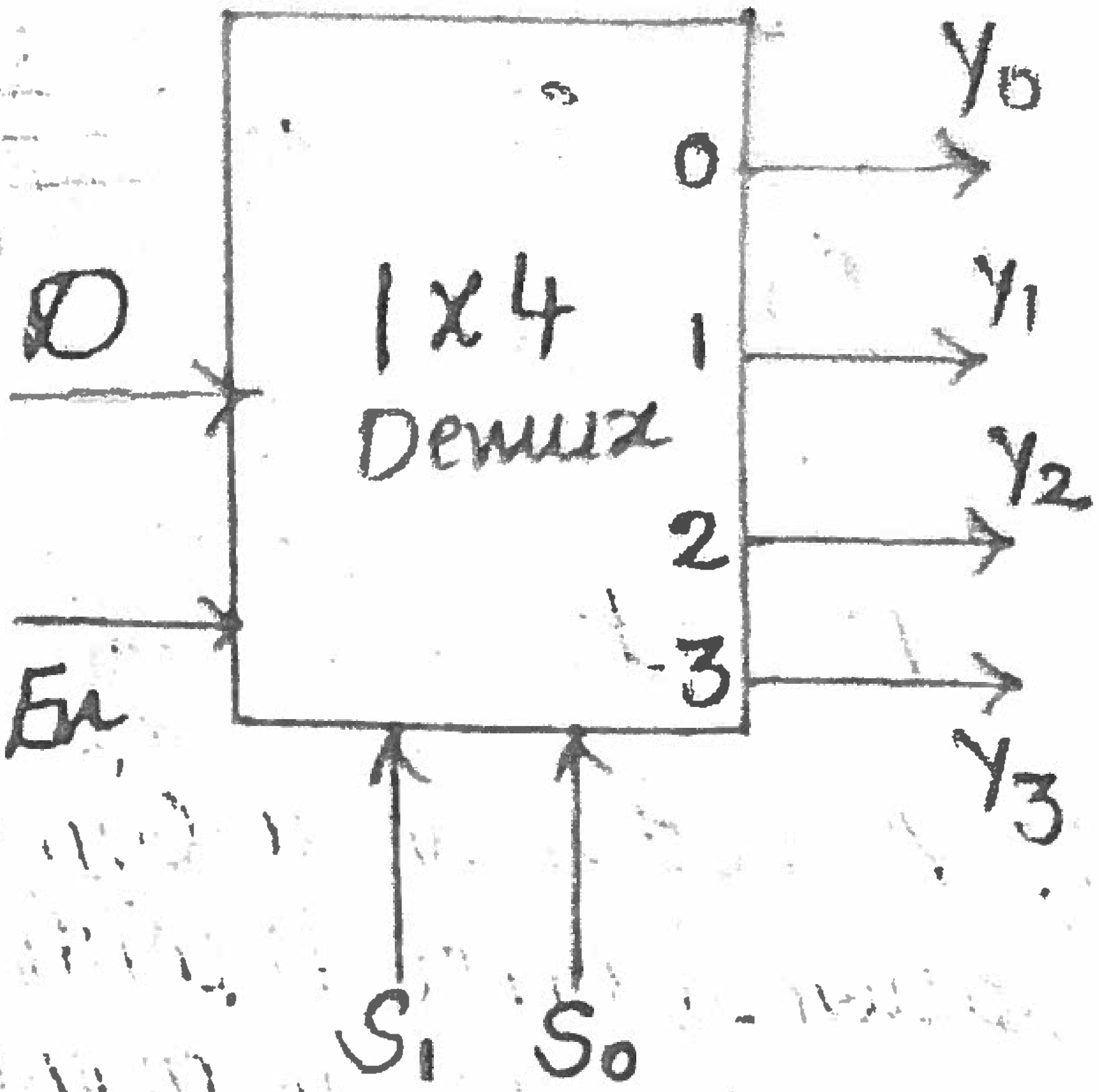
Sol:- Applications :-

1) It is used to connect a single source to multiple destination.

2) Area of Application is Communication System.

3) Serial to Parallel Converter

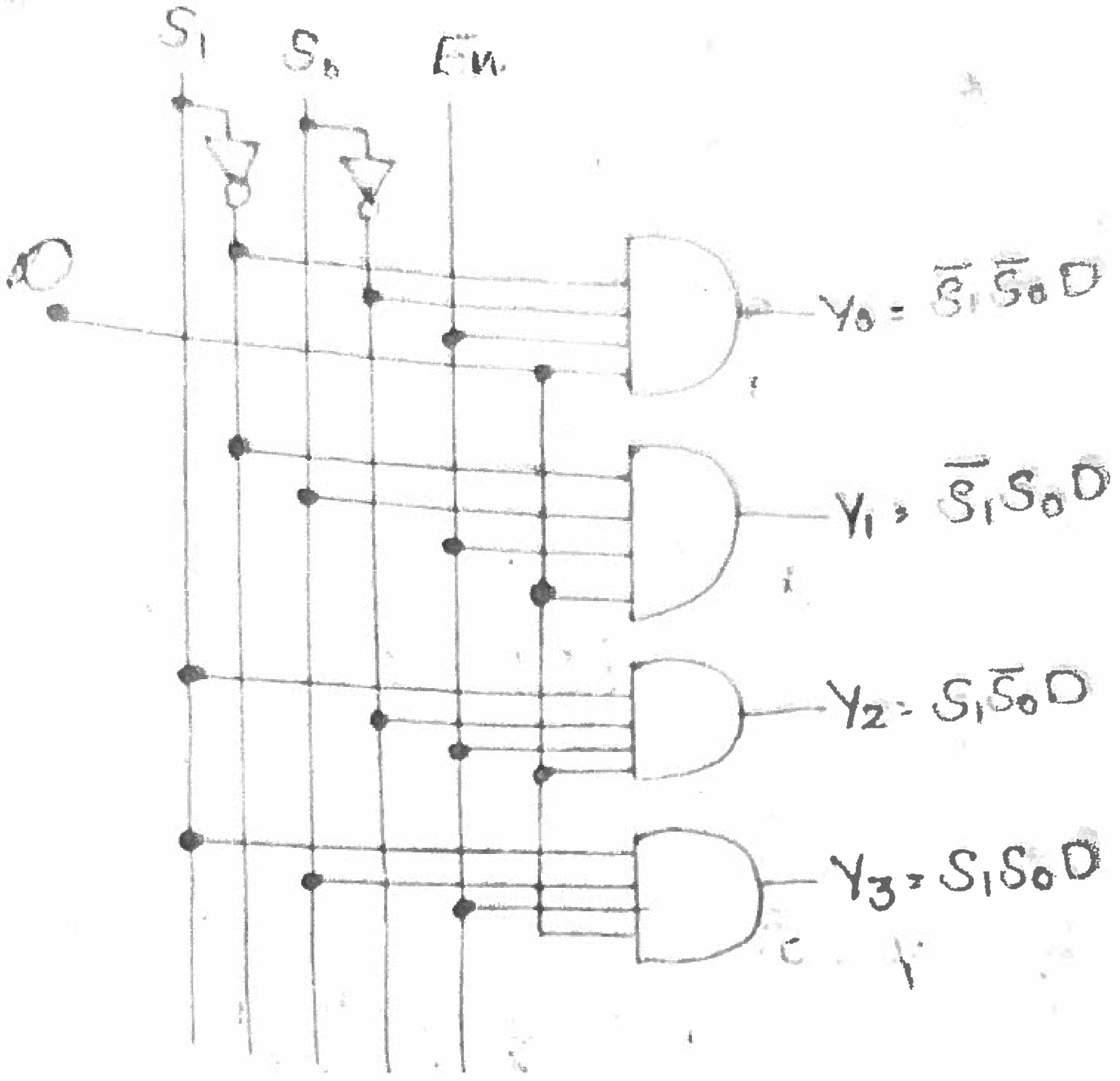
1x4 Demultiplexer



The functional table of 1x4 demux is as follows

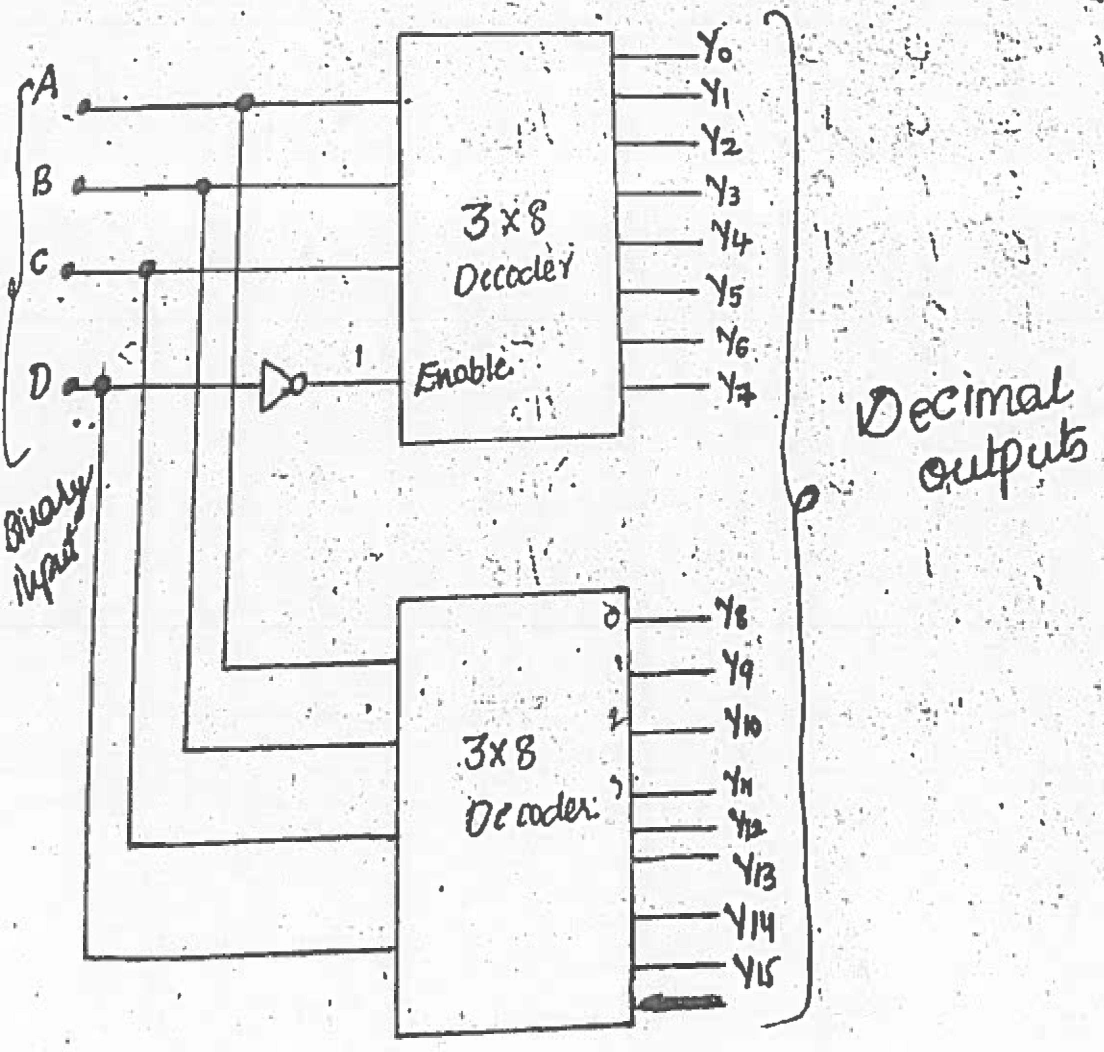
En	S_1	S_0	D	Y_0	Y_1	Y_2	Y_3
0	x	x	x	0	0	0	0
1	0	0	0	0	0	0	0
1	0	0	1	1	0	0	0
1	0	1	0	0	0	0	0
1	0	1	1	0	1	0	0
1	1	0	0	0	0	0	0
1	1	0	1	0	0	1	0
1	1	1	0	0	0	0	0
1	1	1	1	0	0	0	1

diagram



10(b) 4 to 16 Decoder using Two 3 to 8 Decoders (6M)

Decoders with enable inputs can be connected together to form a larger decoder circuit. The circuit arrangement shows 4 to 16 decoder using two 3 to 8 decoder (74138).



the most significant input bit D is connected through an inverter $\bar{E}$ to the upper decoder (Y_0 to Y_7) and directly to E to the lower decoder (Y_8 to Y_{15}). Thus when D=0 (LOW), the upper decoder is enabled and lower decoder is disabled. When D=1 (HIGH), the lower decoder is enabled and upper decoder is disabled.

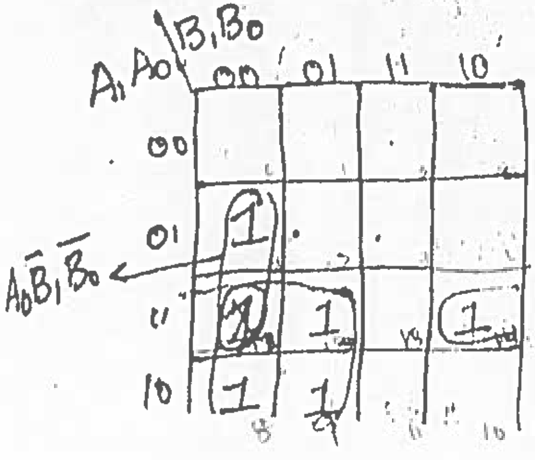
Truth Table.

	Binary Inputs				Decimal output
	D	C	B	A	
Upper decoder	0	0	0	0	Y_0
	0	0	0	1	Y_1
	0	0	1	0	Y_2
	0	0	1	1	Y_3
	0	1	0	0	Y_4
	0	1	0	1	Y_5
	0	1	1	0	Y_6
	0	1	1	1	Y_7
Lower Decoder	1	0	0	0	Y_8
	1	0	0	1	Y_9
	1	0	1	0	Y_{10}
	1	0	1	1	Y_{11}
	1	1	0	0	Y_{12}
	1	1	0	1	Y_{13}
	1	1	1	0	Y_{14}
	1	1	1	1	Y_{15}

11(a) 2-bit comparator: Let the two 2-bit
be $A = A_1 A_0$ & $B = B_1 B_0$. The truth table for
2-bit comparator is as follows. (6M)

Inputs				Outputs		
A_1	A_0	B_1	B_0	$A>B$	$A=B$	$A<B$
0	0	0	0	0	1	0
0	0	0	1	0	0	1
0	0	1	0	0	0	1
0	0	1	1	0	0	1
0	1	0	0	1	0	0
0	1	0	1	0	1	0
0	1	1	0	0	0	1
0	1	1	1	0	0	1
1	0	0	0	1	0	0
1	0	0	1	1	0	0
1	0	1	0	0	1	0
1	0	1	1	0	0	1
1	1	0	0	1	0	0
1	1	0	1	1	0	0
1	1	1	0	1	0	0
1	1	1	1	0	1	0

K-Map for A>B



$$A>B = A_1\bar{B}_1 + A_0\bar{B}_1\bar{B}_0 + A_1A_0\bar{B}_0$$

K-Map for A=B

$A_1A_0 \backslash B_1B_0$	00	01	11	10
00	①			
01		①		
11			①	
10				①

$$A=B \Rightarrow \bar{A}_1\bar{A}_0\bar{B}_1\bar{B}_0 + \bar{A}_1A_0\bar{B}_1B_0 + A_1A_0B_1B_0 + A_1\bar{A}_0B_1\bar{B}_0$$

$$= \bar{A}_1\bar{B}_1(\bar{A}_0\bar{B}_0 + A_0B_0) + A_1B_1(A_0B_0 + \bar{A}_0\bar{B}_0)$$

$$= \bar{A}_1\bar{B}_1(A_0 \odot B_0) + A_1B_1(A_0 \odot B_0)$$

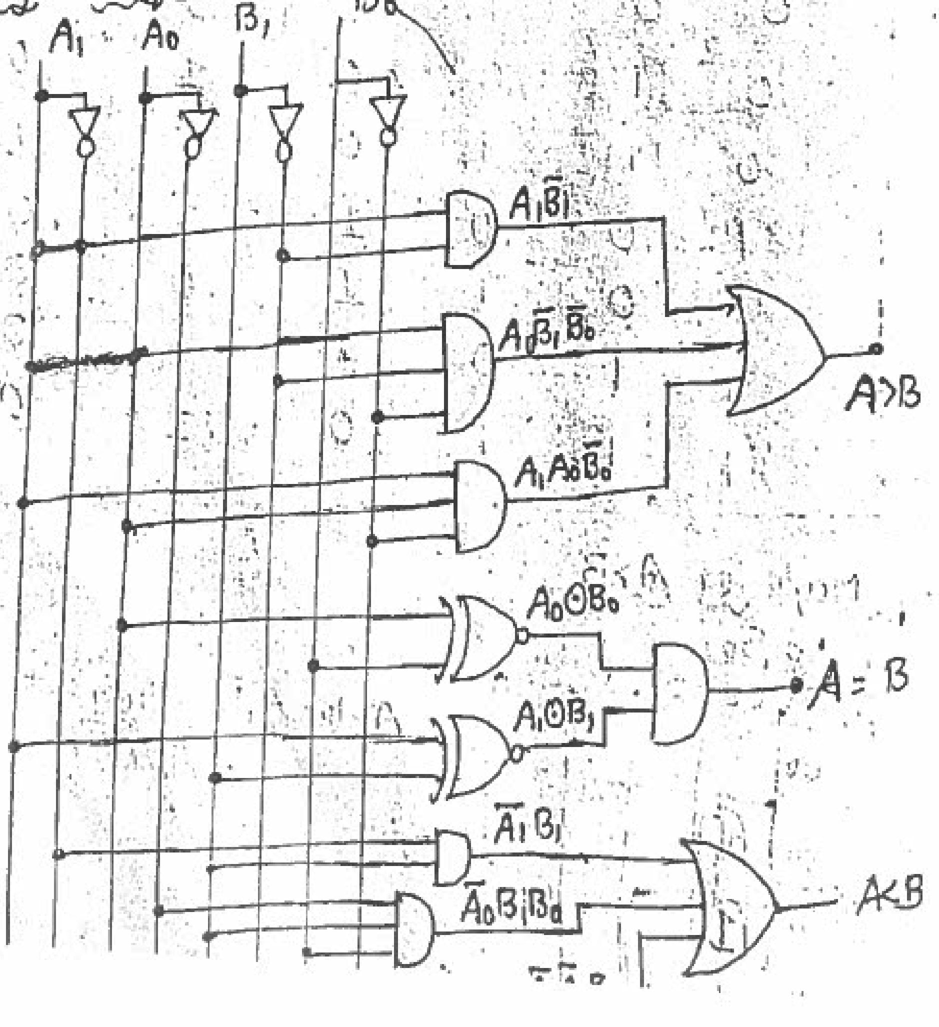
$$= (A_0 \odot B_0)(A_1 \odot B_1)$$

K-Map for A<B

$A_1A_0 \backslash B_1B_0$	00	01	11	10
00		1	1	1
01			1	1
11				
10			1	

$$A<B \Rightarrow \bar{A}_1B_1 + \bar{A}_0B_1B_0 + \bar{A}_1\bar{A}_0B_0$$

Logic diagram:



1(b) Q. Implement the circuit with a PLA having 3 inputs, 3 product terms and two outputs. -(6M)

$F_1 = \sum m(3,5,7)$ $F_2 = \sum m(4,5,7)$

Sol:- K-Map for F_1

True form

A \ BC	00	01	11	10
0			1	
1		1	1	

$F_1 : AC + BC = F_1(T)$

K-Map for F_2

True form

A \ BC	00	01	11	10
0				
1	1	1	1	

$F_2 = A\bar{B} + AC = F_2(T)$

Product term	Inputs A	B	C	outputs F_1	F_2
1 (AC)	1	–	1	1	1
2 (BC)	–	1	1	1	–
3 (A$\bar{B}$)	1	0	–	–	1
				T	T

$F_1(C)$

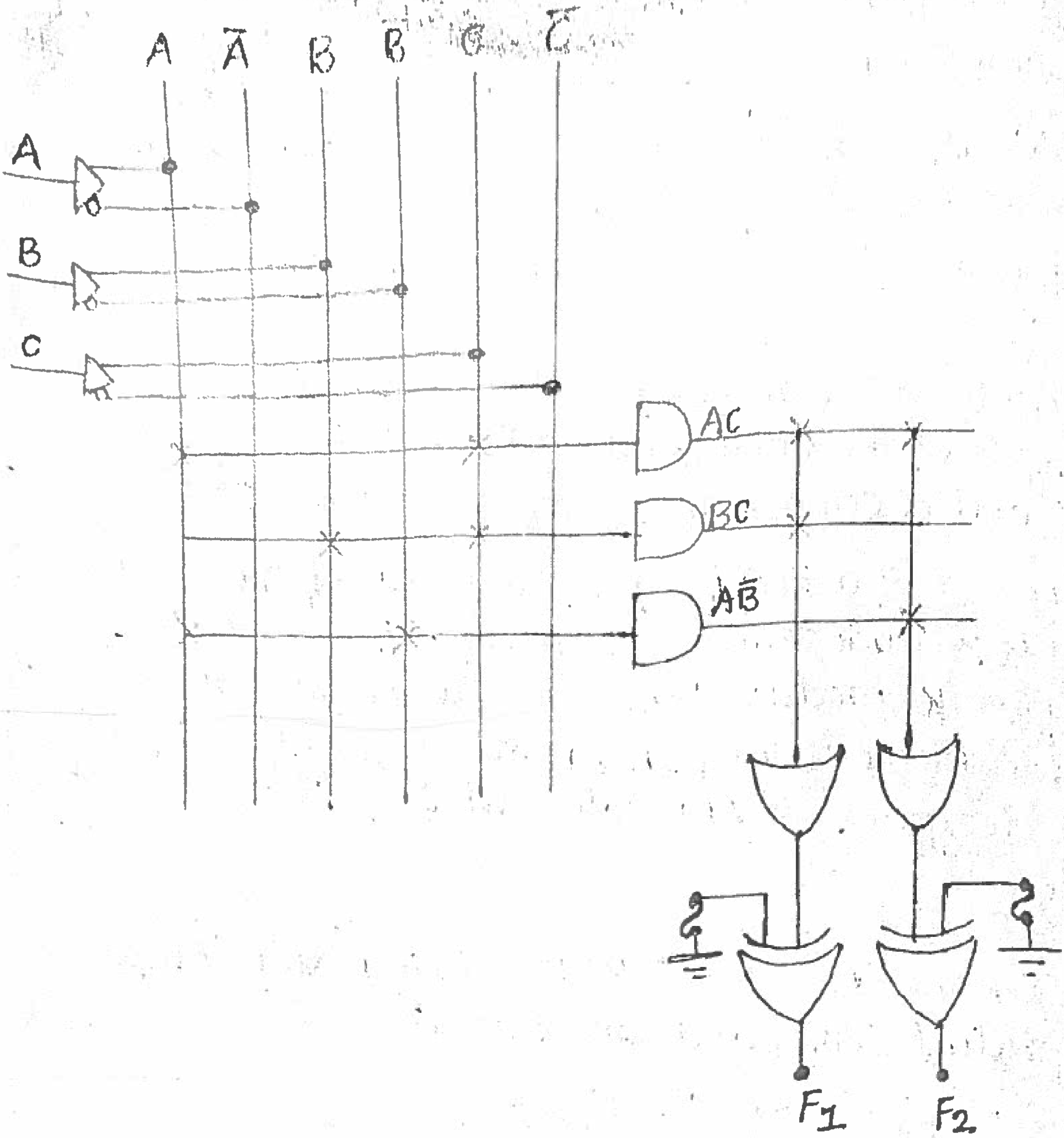
A \ BC	00	01	11	10
0	0	0		0
1	0			0

$\bar{F_1}(C) = \bar{C} + \bar{A}\bar{B}$

A \ BC	00	01	11	10
0	0	0	0	0
1				0

$\bar{F_2}(C) = \bar{A} + B\bar{C}$

$F_2(C) = \overline{\bar{A} + B\bar{C}}$

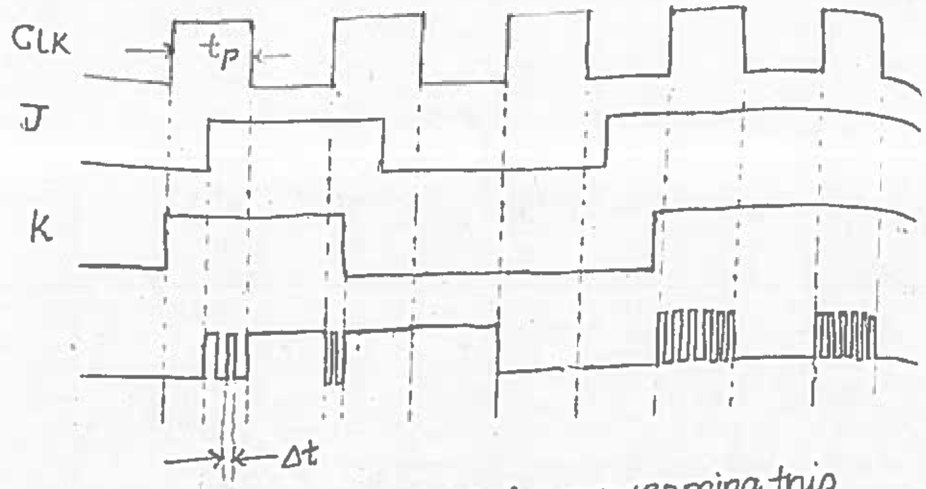

A
Ā
B
B̄
C
C̄
A
B
C
AC
BC
AB̄
F1
F2

12(a) (6M)

Race-Around condition:-

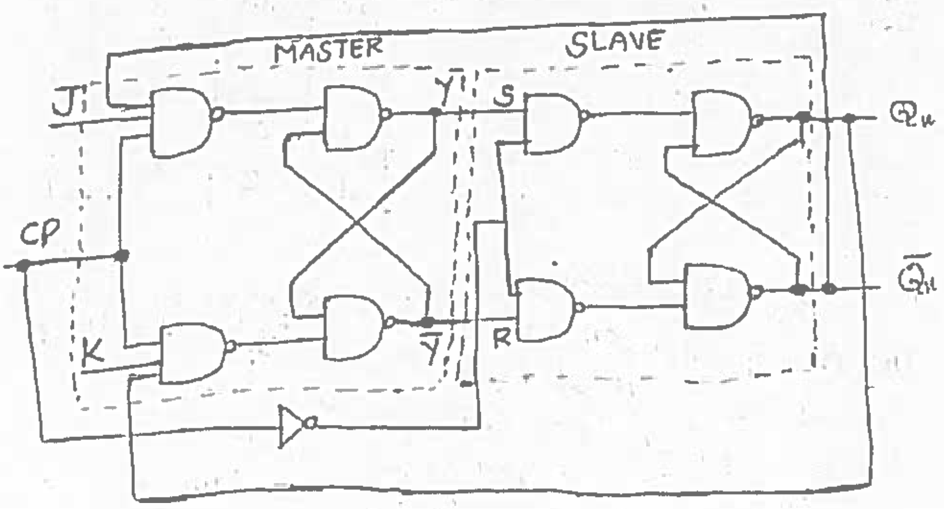
In JK Flip-flop, consider the assignment of excitations J=K=1. If the width of the clock pulse 't_p' is too long the state of the flip-flop will keep on changing from '0' to '1', '1' to '0', '0' to '1' and so on and at the end of the clock pulse, its state will be uncertain. This Phenomenon is known as race around condition.

This condition exists when $t_p \geq \Delta t$. Thus by keeping $t_p < \Delta t$, we can avoid race around condition.



A more practical method for overcoming this difficulty is the use of Master-Slave configuration.

Master-Slave JK Flip-flop:-

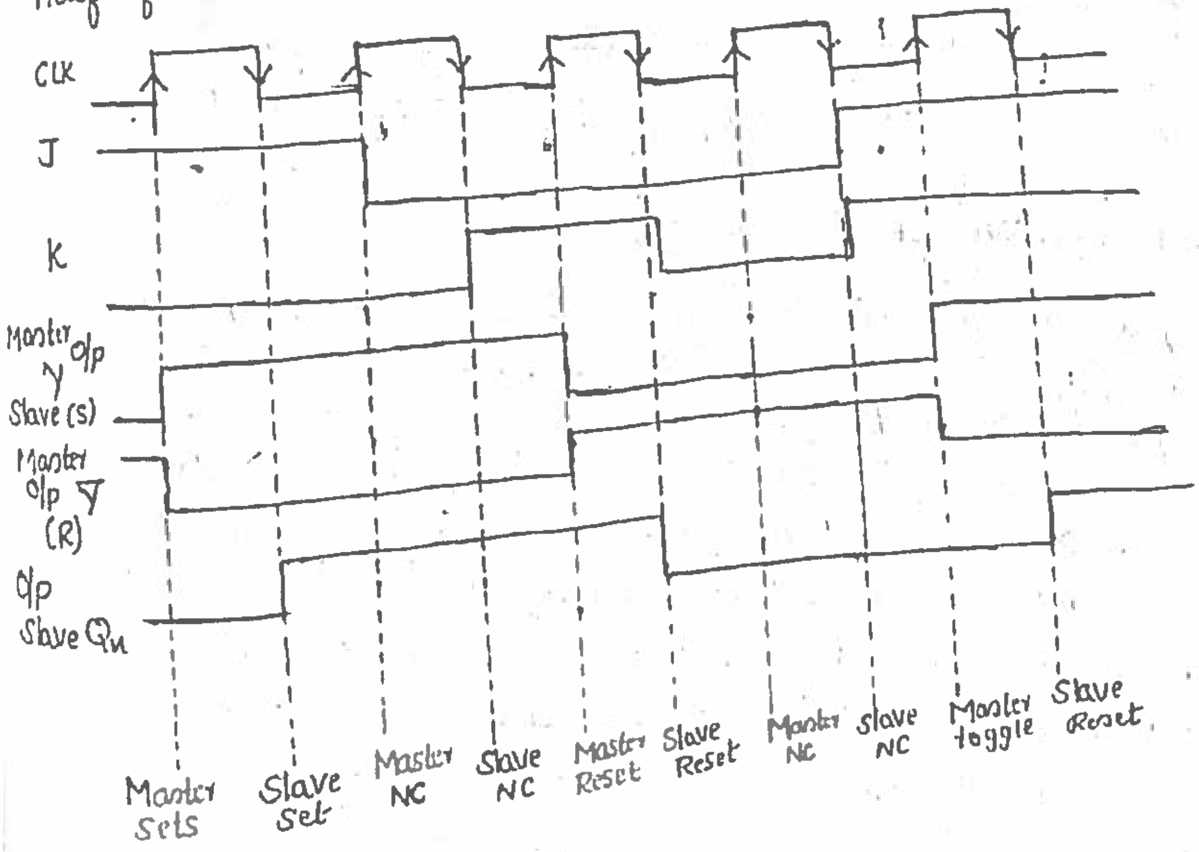


SR FF as a slave.

Working:- When J=1, K=0 the master sets on the positive clock. The high Y output of the master drives the 'S' input of the slave, so at the negative edge of clock slave sets, copying the action of master.

* When J=0, K=1, the master resets on the positive clock. The high $\bar{Y}$ output of the master goes to the 'R' input of the slave. Therefore, at the negative clock slave resets, again copying the action of the master.

* When J=1, K=1 master toggles on the positive clock and slave then copies the output of the master on the negative clock. At this instant, feedback inputs to the master flip-flop are complemented but as it is negative edge half of the clock pulse master FF is inactive.

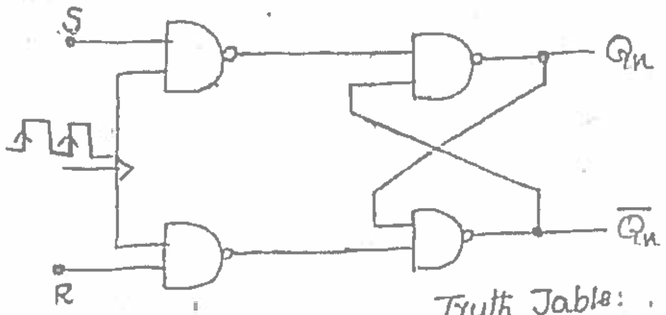


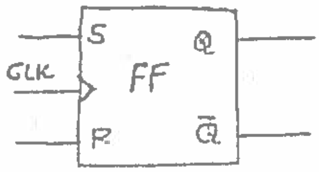
Clk	J	K	Q_n	Y	Q_{n+1}
↑	0	0	0	0	NC
↓	0	0	0	NC	0
↑	0	0	1	1	NC
↓	0	0	1	NC	1
↑	0	1	0	0	NC
↓	0	1	0	NC	0
↑	0	1	1	0	NC
↓	0	1	1	NC	0
↑	1	0	0	1	NC
↓	1	0	0	NC	1
↑	1	0	1	1	NC
↓	1	0	1	NC	1
↑	1	1	0	1	NC
↓	1	1	0	NC	1
↑	1	1	1	0	NC
↓	1	1	1	NC	0

12(b) SR Flip-Flop: (6M)

The following figure shows the positive edge triggered clocked SR FF. The circuit is similar to the SR latch except enable signal is replaced by the clock Pulse (CLK). The circuit output responds to the 'S' & 'R' inputs only at the positive edges of the clock Pulse. At any other instants of time, the SR flip-flop will not respond to the changes in input.

Fig: Positive edge Triggered Flip-flop

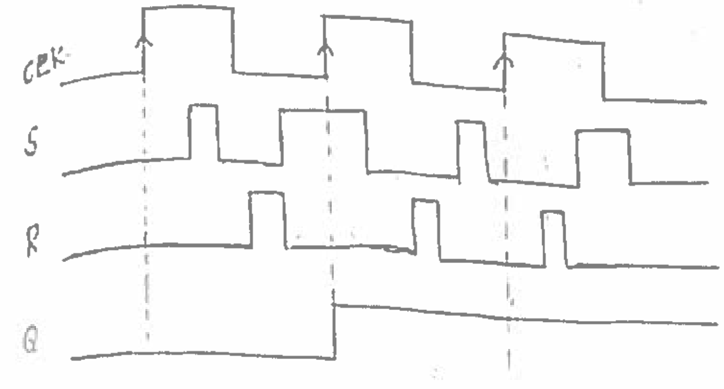




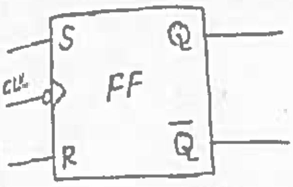
Truth Table:

CLK	S	R	Q_{n+1}
↑	0	0	No change
↑	0	1	Reset
↑	1	0	Set
↑	1	1	Invalid
0	X	X	No change

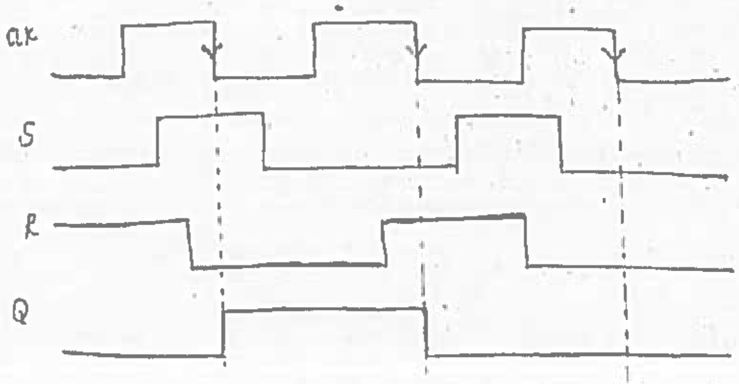
CLK	S	R	Q_n	Q_{n+1}	State
↑	0	0	0	0	No change
↑	0	0	1	1	
↑	0	1	0	0	Reset
↑	0	1	1	0	
↑	1	0	0	1	Set
↑	1	0	1	1	
↑	1	1	0	X	Invalid
↑	1	1	1	X	
0	X	X	0	0	No change
0	X	X	1	1	



In negative edge triggered SR FF, the circuit output responds at the negative edges of the clock pulse. The logic symbol and the truth table of the -ve edge triggered SR flip-flop.



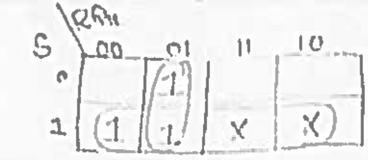
CLK	S	R	Q_{n+1}
↓	0	0	Q_n(NC)
↓	0	1	Reset
↓	1	0	Set
↓	1	1	Invalid
0	X	X	No change



Characteristic Equation of SR Flip-Flop:-

The characteristic equation of a flip-flop is the equation expressing the next state of a flip-flop in terms of its present state and present excitations.

K-Map for Q_{n+1}



S \ RQ_n	00	01	11	10
0		1		
1	1	1	X	X

$$\Rightarrow \boxed{Q_{n+1} = S + \bar{R}Q_n}$$

Excitation Table: For the design of sequential circuits we should know the excitation table of flip-flops. The excitation table of the flip-flop can be obtained from its truth table. It indicates the inputs required to be applied to the flip-flop to take it from present state to next state.

Excitation Table of SR flip-flop is as follows:

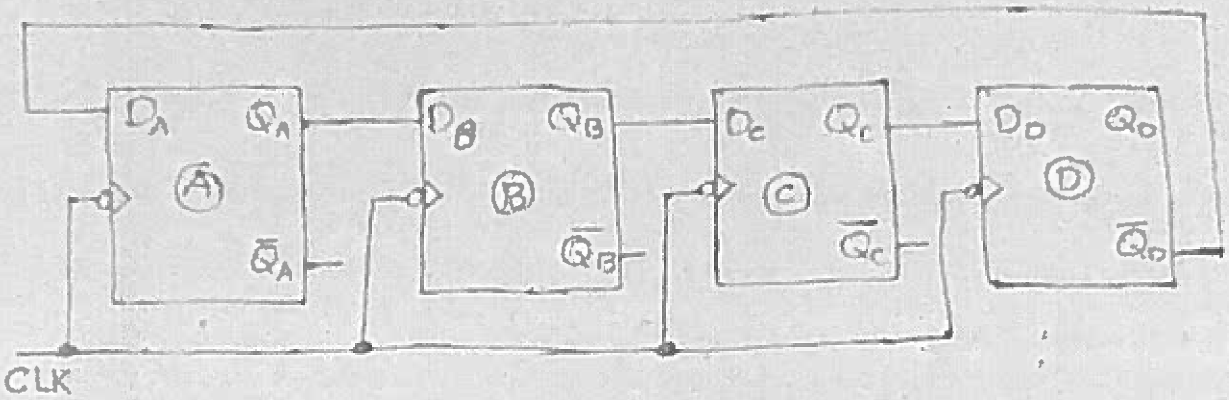
PS	NS	Required inputs	
Q_n	Q_{n+1}	S	R
0	0	0	X
0	1	1	0
1	0	0	1
1	1	X	0

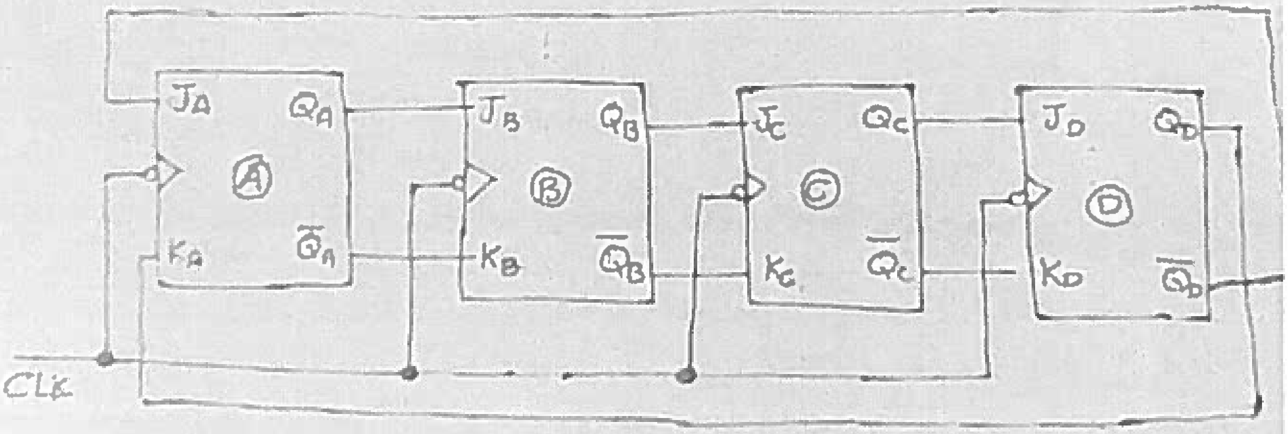
13(a)

JOHNSON COUNTER (TWISTED RING COUNTER) -(6M)

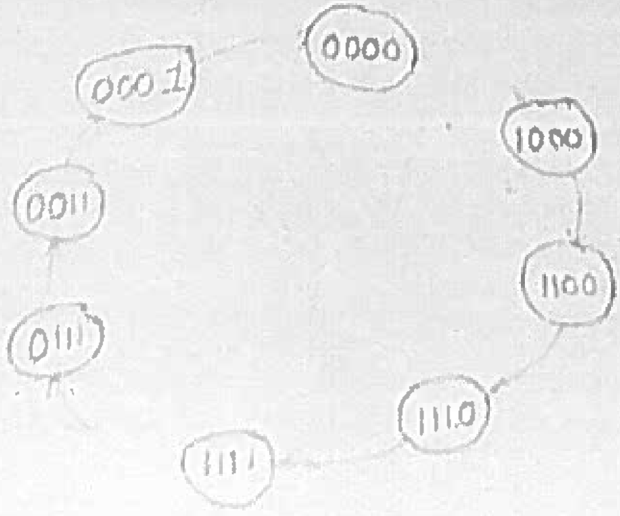
This counter is obtained from a ring counter by providing feedback from the inverted output of the last FF to the 'D' input of the first FF.

* The Q input of each stage is connected to the D input of the next stage, but the $\overline{Q}$ output of the last stage is connected to the 'D' input of the first stage, therefore the name is twisted counter.





The state diagram of a johnson counter.

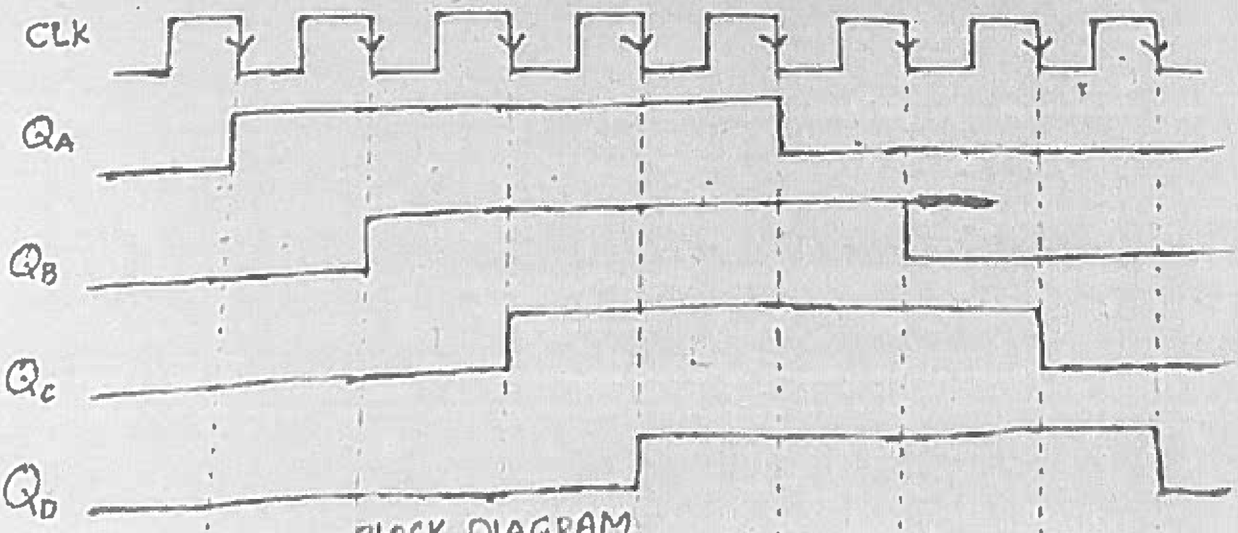


Sequence Table:

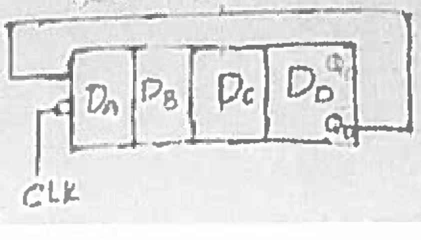
Q_A	Q_B	Q_C	Q_D	After CLK pulses
0	0	0	0	0
1	0	0	0	1
1	1	0	0	2
1	1	1	0	3
1	1	1	1	4
0	1	1	1	5
0	0	1	1	6
0	0	0	1	7
0	0	0	0	8
1	0	0	0	9

Let ~~the~~ initially all the FF's are Reset, i.e., the state of the counter is 0000. After each clock pulse, the level of Q_A is shifted to Q_B, The level of Q_B is shifted to Q_C, The level of Q_C is shifted to Q_D, the level of $\overline{Q_D}$ is shifted to Q_A.

* A Johnson Counter of n-FF's will have 2n unique states and count upto 2n pulses. So it is a MOD-2n counter. It is more economical than ring counter but less economical than ripple counter.



BLOCK DIAGRAM

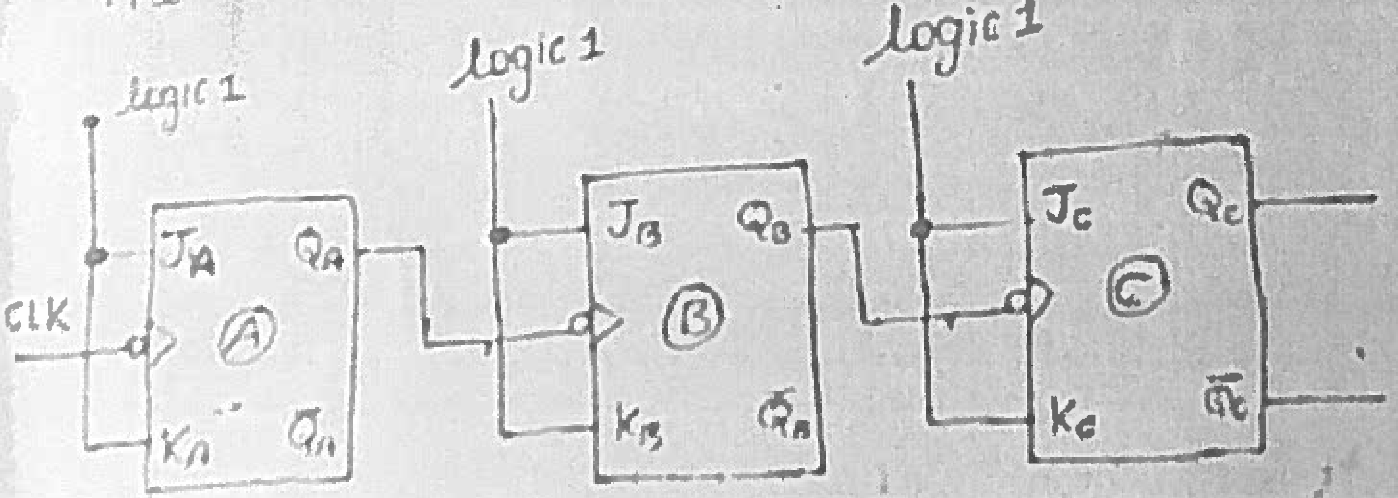


13(b)

Asynchronous counters:- —(6M)

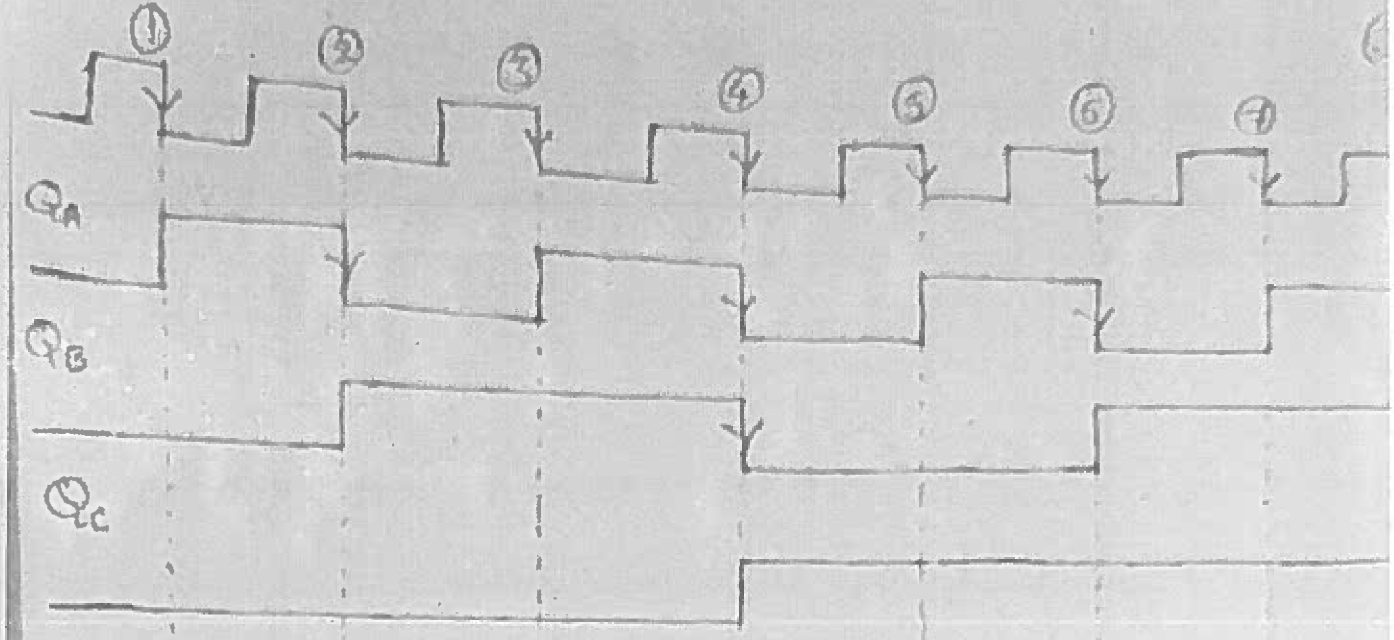
Three-bit Ripple up-counter using Negative edge triggered FF's:

+ The three-bit up-counter consists of 3 FF's and this counter counts in the order 0,1,2,3,4,5,6,7,0. The external clk pulse is applied to one of the FF and its output is considered as the LSB.



a) Logic diagram

The timing diagram for the 3-bit up counter using -ve edge triggered FF's is as follows:



Clock pulse	Q_C	Q_B	Q_A	Decimal
2	~~0~~	~~0~~	~~1~~	~~1~~
Initially	0	0	0	0
1st (↓)	0	0	1	1
2nd (↓)	0	1	0	2
3rd (↓)	0	1	1	3
4th (↓)	1	0	0	4
5th (↓).	1	0	1	5
6th (↓)	1	1	0,	6
7th (↓).	1	1.	1	7
8th (↓)	0	0	0	0

14) a) Structure of VHDL Program :

(6 M)

Every VHDL program consists of at least one entity/architecture pair. In a large design, you will typically write many entity/architecture pairs and connect them together to form a complete circuit. An entity declaration describes the circuit as it appears from the "outside" - from the perspective of its input and output interfaces. The second part of a minimal VHDL design description is the architecture declaration.

ENTITY :

ENTITY is the list with specifications of all input and output pins of the circuit. Its syntax is shown below :

```
ENTITY name IS

PORT (

port_name : signal_mode signal_type;

port_name : signal_mode signal_type;

...

);

END name;
```

The mode of the signal used may be IN, OUT, INOUT or BUFFER. IN and OUT are unidirectional pins, while INOUT is bidirectional. BUFFER, is used when the output signal is used internally in the design. The type of the signal may be BIT, STD_LOGIC, INTEGER, etc. The name of the entity should be not use the VHDL reserved words.

Let us consider the NAND gate, Its ENTITY may be declared as :

```
ENTITY nand_Two IS

PORT (a, b : IN BIT;

x : OUT BIT
```

```
);
END nand_Two;
```

ARCHITECTURE :

The ARCHITECTURE is the description of how the circuit of design works. Its syntax is as follows,

```
ARCHITECTURE architecture_name OF entity_name IS
[declarations]
BEGIN
(code)
END architecture_name;
```

As shown in above syntax, architecture has two parts, 1) A declarative part where signals and constants are declared and 2) The code part. The name of architecture is any name except VHDL reserved words.

14 b)Design of 4 to 1 Multiplexer using CASE Statement (Behavior Modeling Style) (6M)

```
library IEEE;
use IEEE.STD_LOGIC_1164.all;

entity multiplexer_case is
    port(
       din : in STD_LOGIC_VECTOR(3 downto 0);
       sel : in STD_LOGIC_VECTOR(1 downto 0);
       dout : out STD_LOGIC
       );
end multiplexer_case;


architecture multiplexer_case_arc of multiplexer_case is
begin

    mux : process (din,sel) is
    begin
       case sel is
```

```
        when "00" => dout <= din(3);
        when "01" => dout <= din(2);
        when "10" => dout <= din(1);
        when others => dout <= din(0);
      end case;
    end process mux;

end multiplexer_case_arc;
```

15 .b)Explain about Package in VHDL? (6M)

- Groups of procedures and functions that are related can be aggregated into a module that is called package.
- A package can be shared across many VHDL models (type definitions, functions, procedures, etc)
- A package can also contains user defined data types and constants.
- When working with a large design project consisting of many small VHDL programs, it is convenient to have common procedures and functions in separate packages.

Package declaration

It contains interface or specifications of the functions and procedures that are available in the package. It gives idea about – list of functions and procedures, the input parameters, the type of input parameters,

the output value, the type of the output value etc.

Package declaration Syntax:

packagepackage-name is

package-item-declarations_ these may be:

subprogram declarations / type declarations / subtype declarations / constant declarations /

signal declarations / variable declarations / file declarations / alias declarations / component declaration / attribute declarations / attribute specifications /disconnection specifications

end packagepackage-name;

Package Body

It basically contains the code that implements the subprograms

Package Body Syntax:

package body package-name is

subprogram bodies / complete constant declarations / subprogram declarations / type and subtype declarations / file and alias declarations / use clauses

end package bodypackage-name;

Library in VHDL

Design units such as packages, architectures and Entities can be compiled into a library. Libraries are generally implemented as directories on a computing system.
Library is referenced by its logical name and there exists a mapping of logical name into the physical name of the system (which is the directory names on the computing system).
The mapping is maintained by the system.
Just like variables and signals, library must be declared before we can use it.
Syntax for Library declaration:

```
LIBRARY logical_library_name_1, logical_library_name2, ...;
```

In the VHDL language, the libraries STD and WORK are implicitly declared in the source code.
Library STD contains the standard packages with VHDL distribution.
The WORK library refers to the current working directory.
There are other different libraries used for different tools used.
Other libraries such as math and other misc. are often supplied with the tool

15) a)Explain Data flow modeling in VHDL. (6 M)

In this modeling style the flow of data is monitored from input to output inside an entity.
Data Flow Modeling Style works on Concurrent Executions.
Concurrent Signal Assignment (<=), With Select Assignment, When Else Assignment
Generate Expressions are generally used in Data Flow Modeling Style.
Programs:

VHDL code for AND-OR-INVERTER (AOI) in dataflow modeling style?

```
Library IEEE;
Use IEEE.STD_LOGIC_1164.all;
Entity AOI is
Port(A,B,C,D: std_logic;
        Y: out_logic);
End AOI;
Architecture dataflow1 of AOI is
Begin
Y=not((A and B) or (c and D));
End dataflow1;
```

Semester End Supplementary Examination, April/May, 2022

Degree	B. Tech. (U. G.)	**Program**	CSE, CSE (AI & ML) & CS (DS)	**Academic Year**	2021 - 2022
Course Code	20CS305	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	III
Course	**COMPUTER ORGANIZATION**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the most commonly used binary code?	20CS305.1	L1
2	What is binary adder?	20CS305.2	L1
3	What is control memory address?	20CS305.3	L1
4	How do you handle divide overflow?	20CS305.4	L2
5	Is CDROM an auxiliary memory?	20CS305.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain floating point representation of numbers in the ALU.	6M	20CS305.1	L2
6 (b)	Convert the number 7425 into octal, hexadecimal and BCD.	6M	20CS305.1	L2
	OR			
7 (a)	Distinguish between error detection and correction codes.	6M	20CS305.1	L2
7 (b)	State the differences between encoder and multiplexer. Mention the role of these components in the design of computers	6M	20CS305.1	L2
8 (a)	What is a micro-operation? Discuss the four different types of micro-operations.	5M	20CS305.2	L2
8 (b)	Explain the different phases of an instruction cycle. What happens in case an instruction has some memory operands?	7M	20CS305.2	L2
	OR			
9 (a)	What is register transfer language? Describe the basic symbols used in register transfer.	5M	20CS305.2	L2
9 (b)	Discuss with example any four memory reference instructions.	7M	20CS305.2	L2
10 (a)	Explain about control memory in a micro programmed control organization.	8M	20CS305.3	L2
10 (b)	What do you mean by Complex Instruction Set Computer (CISC)? Discuss relative advantages and disadvantages of such instruction set design.	4M	20CS305.3	L2
	OR			
11 (a)	Explain the basic computer instruction formats.	5M	20CS305.3	L2
11 (b)	What do you mean by Addressing modes? Explain the following addressing modes: i) Index Addressing mode ii) Immediate Addressing mode iii) Relative Addressing mode iv) Direct Addressing mode.	7M	20CS305.3	L2
12 (a)	Discuss Booth multiplication algorithm.	6M	20CS305.4	L2
12 (b)	Explain the functional units and their data flow in a hardware implementation that performs addition and subtraction of signed-magnitude numbers.	6M	20CS305.4	L2
	OR			
13 (a)	Draw a Flow chart which explains multiplication of TWO signed magnitude fixed point numbers.	6M	20CS305.4	L2
13 (b)	Draw a flow chart for adding and subtracting two fixed point binary numbers where negative numbers signed 1's complement presentation.	6M	20CS305.4	L2
14 (a)	What is the significance of cache memory and write about direct and associative mapping techniques.	7M	20CS305.5	L2
14 (b)	Distinguish between synchronous and asynchronous data transfer.	5M	20CS305.5	L2
	OR			
15 (a)	With the help of a block diagram explain the concept of DMA controller.	6M	20CS305.5	L2
15 (b)	What is virtual memory? With a neat block diagram explain the virtual memory address translation.	6M	20CS305.5	L2

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N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

III Semester End Supplementary Exam

April 2022

Computer Organization (20CS305)

(Common to CSE, CSE(AI &ML),CSE(DS))

Part A (Short Answer Question 5X2 =10 Marks)

1. What is the most commonly used binary code?
ASCII-1 M+ Explain purpose of ASCII – 1M
ASCII (American Standard Code for Information Interchange) is the most commonly used binary code. In an ASCII file, each alphabetic, numeric, or special character is represented with a 7-bit binary number (a string of seven 0s or 1s). ASCII was soon expanded to an 8-bit system that has 256 code points.
ASCII codes represent text in computers, telecommunications equipment, and other devices. Most modern character-encoding schemes are based on ASCII, although they support many additional characters.

2. What is binary adder?
Definition of binary adder with explanation and circuit-2 M
A binary adder is a digital circuit that produces the arithmetic sum of two binary numbers. It can be constructed with full adders connected in cascading form with the output carry from each full adder connected to the input carry of the next full adder in chain. Addition of n-bit numbers requires a chain of n full adders or chain of 1 half-adder and n-1 full adders.

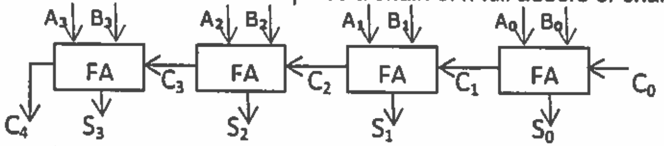


Fig: 4-bit binary adder

3. What is control memory address?
Control memory address purpose— 2 M
The control memory address register specifies the address of the microinstruction. The microinstruction contains a control word that specifies one or more microoperations for the data processor. The next address is computed in the next address generator also called microprogram sequencer. Typical functions of a microprogram sequencer are incrementing the control address register by one, loading into the control address register an address from control memory, transferring an external address, or loading an initial address to start the control operations.

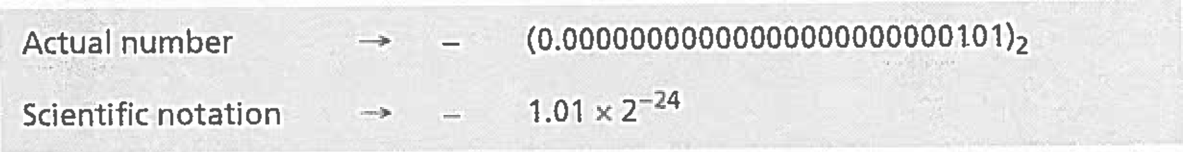
4. How do you handle divide overflow?
Handling for divide overflow — 2 M
In some computers it is the responsibility of the programmers to check if DVF if set after each divide instruction. They can branch to a subroutine that takes a corrective measure. In some older computers divide overflow stopped the computer. But it is time consuming. The procedure in most computers is to provide an interrupt request when DVF is set.The interrupt causes the computer to suspend the current program and branch to a service routine to take a corrective measure. The most common corrective measure is to remove the program and type an error message explaining the reason why the program could not be completed. It is then the responsibility of the user who wrote the program to rescale the data or take any other corrective measure. The best way to avoid a divide overflow is to use floating point data.

5. Is CDROM an auxiliary memory?
CDROM is auxiliary memory and explanation of auxiliary memory-2M

highest-capacity and slowest-access storage in a computer system.In these memories, programs and data are kept for long-term storage or when not in immediate use. Examples of auxiliary memories are magnetic tapes, magnetic disks, floppy,CD-ROM

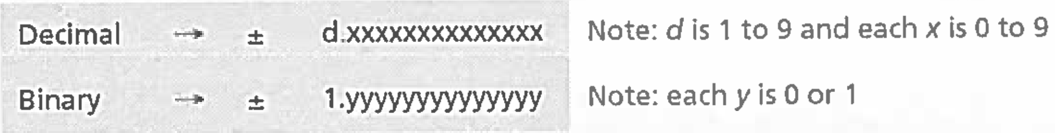
Part B (Long Answer Questions 5X12 = 60 Marks)

6 (a) Explain floating point representation of numbers in the ALU
Scientific notation with example-3M+Normalized notation with example-3M
Floating Point Notation is a way to represent very large or very small numbers precisely using scientific notation in binary.
The floating-point notation has two types of notation
Scientific notation and Normalized notation
Scientific notation – Method of representing binary numbers into a x b^e form where a is real number, b is base and e is exponent.
Scientific notation is further converted into floating-point notation because floating-point notation only accepts scientific notation.
For example:-x = 376.4423 (its not scientific notation)
Number in scientific = 36.4423 x 10^1 or 3.64423 x 10^2
To represent the number $-(0.000000000000000000000000101)_2$ in floating point
Keeping only one digit to the left of the decimal point we get



Actual number	→	−	$(0.000000000000000000000000101)_2$
Scientific notation	→	−	1.01×2^{-24}

Normalized notation- To make the fixed part of the representation uniform, both the scientific method (for the decimal system) and the floating-point method (for the binary system) use only one non-zero digit on the left of the decimal point. This is called normalization. In the decimal system this digit can be 1 to 9, while in the binary system it can only be 1. In the following, d is a non-zero digit, x is a digit, and y is either 0 or 1.



Decimal	→	±	d.xxxxxxxxxxxxxx	Note: *d* is 1 to 9 and each *x* is 0 to 9
Binary	→	±	1.yyyyyyyyyyyyyy	Note: each *y* is 0 or 1

0 cannot be represented or normalized because the representation set starts from 0.1, so how can we normalize zero. It's not possible.If the most significant bit of mantissa is a non zero, then such a representation is called normalized floating-point.

6 (b) Convert the number 7425 into octal, hexadecimal and BCD.
Octal conversion-2 M + Hexadecimal conversion of 7425 - 2 Marks + BCD Conversion-2M
Base for octal is 8, hexadecimal is 16.
Divide 7425 by the base number keeping notice of the quotient and the remainder.
Continue dividing the quotient by base until you get a quotient of zero.
Then just write out the remainders in the reverse order to get the equivalent of decimal number 7425
(i) Octal of 7425 is obtained by dividing 7425 with 8

7425 / 8 = 928 with remainder 1
928 / 8 = 116 with remainder 0
116 / 8 = 14 with remainder 4
14 / 8 = 1 with remainder 6
1 / 8 = 0 with remainder 1
Octal of 7425 is $(16401)_8$

(ii) Hexadecimal of 7425 is obtained by dividing 7425 with 16

7425 / 16 = 464 with remainder 1
464 / 16 = 29 with remainder 0
29 / 16 = 1 with remainder 13 (D in hexadecimal)
1 / 16 = 0 with remainder 1
Hexadecimal of 7425 is $(1D01)_{16}$
(iii) BCD of 7425 is 0111 0100 0010 0101

7 (a) Distinguish between error detection and correction codes

Error detection and error correction differences along with parity bit, CRC etc– 6M

Error Detection is the detection of errors caused by noise or other impairments during transmission from the transmitter to the receiver.
Error correction is the detection of errors and reconstruction of the original, error-free data.
Parity Bit: A parity bit is a bit that is added to a group of source bits to ensure that the number of set bits (i.e., bits with value 1) in the outcome is even or odd. It is a very simple scheme that can be used to detect single or any other odd number (i.e., three, five, etc.) of errors in the output. An even number of flipped bits will make the parity bit appear correct even though the data is erroneous.
For example, if each of a series of m-bit "words" has a parity bit added, showing whether there were an odd or even number of ones in that word, any word with a single error in it will be detected. It will not be known where in the word the error is, however. If, in addition, after each stream of n words a parity sum is sent, each bit of which shows whether there were an odd or even number of ones at that bit-position sent in the most recent group, the exact position of the error can be determined and the error corrected.
Any error-correcting code can be used for error detection. A code with minimum Hamming distance, d, can detect up to d − 1 errors in a code word. Using minimum-distance-based error-correcting codes for error detection can be suitable if a strict limit on the minimum number of errors to be detected is desired.
Codes with minimum Hamming distance d = 2 are degenerate cases of error-correcting codes, and can be used to detect single errors. The parity bit is an example of a single-error-detecting code.

7 (b) State the differences between encoder and multiplexer. Mention the role of these components in the design of computers
Atleast 3 differences between encoder and multiplexer-4 M+Role of encoder and multiplexer in computer design-2M

Encoder	Multiplexer
An encoder is a combinational circuit that converts binary information from "n" input lines to a max of 2^n unique output lines. n→2^n	A multiplexer is a combinational circuit that selects one of several digital input signals and send those inputs to the final output. 2^n→1
The encoder does not have any selection input lines	A multiplexer of 2^n inputs allows n selection lines to select the number of input lines to send the final output
There is no need of select input line and limited data can be sent thereby	The select lines determine which input is connected to the output and also increase the amount of data that can be sent over a network within a certain time.

An encoder is a device that can be used to change a signal or data into a specific code. The encoder is used for compression of data. The encoder will convert the information from one format to another format i.e like electrical signals to counters. The feedback signal of the encoder will determine the position, count, speed and direction. The control devices are used to send the command to a particular function.
Examples of encoders are priority encoder, decimal to BCD encoder, octal to binary encoder. The code may be used for any purposes like, for compression of information required for transmission and storage.
In telephone networks, multiplexers are used where multiple audio signals are integrated on a single line of transmission.
Multiplexers are also used as parallel to serial data converters.
A common application of multiplexing occurs when several embedded system devices share a single transmission line or bus line while communicating with the device. Each device in succession has a brief time to send and receive the data.
8 (a) What is a micro-operation? Discuss the four different types of microoperation.
Microoperation-1 M+Four types of microoperation-Arithmetic, Logical, Shift and Register transfer microoperation -1 Mark for each(4M)
Microoperation: The operations executed on data stored in registers is called microoperation.
Arithmetic microoperations- Perform arithmetic operations on numeric data in registers
The basic arithmetic microoperations are: addition, subtraction, increment and decrement.
$R3 \leftarrow R1 + R2$. It states that contents of registers R1 and R2 are added and sum is transferred to register R3
Subtraction Microoperation: $R3 \leftarrow R1 - R2$ or : $R3 \leftarrow R1 + \overline{R2} + 1$
Logical microoperations- Perform bit by bit operations on nonnumeric data in registers
Logic microoperations are: Clear, Exclusive Or, AND, OR, NAND etc
Shift microoperations are microoperations that are used for serial transfer of data. It shifts data either to left or right
Example : Logical Shift, Arithmetic Shift, Circular Shift (either left or right)
Register Transfer : The symbolic notation used to describe the micro-operation transfers amongst registers is called Register transfer language. Example : $R3 \leftarrow R1$ and Read : $DR \leftarrow M[AR]$, Write : $M[AR] \leftarrow R1$

8 (b) Explain the different phases of an instruction cycle. What happens in case an instruction has memory operands?
Fetch, Decode and Execute→Instruction cycle-4M+If instruction has memory operands explain what happens-3M
Different phases of instruction cycle are as follows:
1. Fetch the instruction from memory
2. Decode the instruction

3. Read the effective address if the instruction has an indirect address

4. Execute the instruction.

Upon completion of step 4, the control goes back to step 1 to fetch, decode and execute the next instruction. This process continues indefinitely unless a HALT instruction is encountered.

Fetch and Decode : Initially, the program counter PC is loaded with the address of the first instruction in the program. The sequence counter SC is cleared to 0, providing a decoded timing signal T_0. After each clock pulse, SC is incremented by one, so that the timing signals go through a sequence T_0, T_1, T_2, and so on. The microoperations for the fetch and decode phases can be specified by the following register transfer statements.

T_0: AR←PC
T_1: IR←M[AR], PC←PC+1
T_2: $D_0 \ldots D_7$←Decode IR(12-14), AR← IR(0-11), I ← IR(15)

Decoder output D_7, is equal to 1 if the operation code is equal to binary 111. If $D_7 = 0$ then it is a memory-reference instruction and if I =1 then it means it's a indirect address which is achieved using AR←M[AR].

This address is used during the memory read operation. The word at the address given by AR is read from memory and placed on the common bus. The LD input of AR is then enabled to receive the indirect address that resided in the 12 least significant bits of the memory word. The three instruction types are subdivided into four separate paths. The selected operation is activated with the clock transition associated with timing signal T_3.This can be symbolized as follows:

D_7 IT_3: AR←M[AR]
D_7 $I'T_3$: Nothing

When a memory-reference instruction with I = 0 is encountered, there is no need to do anything. The sequence counter SC must be incremented when $D_7'T_3$, = 1, so that the execution of the memory-reference instruction can be continued with timing variable T_4.

9 (a) What is a register transfer language? Describe the basic symbols used in register transfer

Register transfer language – 2 M + Basic symbols used in register transfer – 3 M

Register Transfer Language: Shortly known as RTL. The symbolic notation used to describe the micro-operation transfers amongst registers is called Register transfer language.

Basic symbols used in register transfer are as follows :

Basic Symbols for Register Transfers

Symbol	Description	Examples
Letters (and numerals)	Denotes a register	*MAR, R2*
Parentheses ( )	Denotes a part of a register	*R2(0–7), R2(L)*
Arrow ←	Denotes transfer of information	*R2 ← R1*
Comma ,	Separates two microoperations	*R2 ← R1, R1 ← R2*

Example : R2 ←R1 means transfer the content of register R1 to register R2

Read : DR←M[AR], Write : M[AR] ←R1

9 (b) Discuss with examples any four memory reference instructions

Mention any four memory-reference instructions among AND, ADD, LDA, STA, BUN, BSA and ISZ – 3M+ Example for 4 memory reference instructions-1 M for each example Total—4M

For example: AND-Performs AND logic operation on bits in AC and memory word given by the effective address

LDA-Load to AC-Transfers the memory word specified by the effective address to AC.

STA- Store Accumulator-Store the content of AC into the memory word specified by effective address. It is represented by D_3.

AND requires operation decoder D_0 and is symbolically represented as

D_0T_4 : DR← M[AR]
D_0T_5 : AC←AC ∧ M[AR], SC←0

ADD requires operation decoder D_1 and is symbolically represented as

D_1T_4 : DR← M[AR]
D_1T_5 :AC←AC + M[AR], E←C_{out}, SC←0

LDA requires operation decoder D_2 and is symbolically represented as

D_2T_4: DR ← M[AR]

D_2T_5: AC←DR, SC←0

STA requires just one microoperation since the output of AC is applied to the bus and the data input of memory is connected to the bus.

D_3T_4:M[AR] ←AC, SC←0

10 (a) Explain about control memory in a micro programmed control organization.

Control memory definition -2M + Purpose of control memory-3M+Diagram of microprogrammed control organization-2M

The control memory address register specifies the address of the microinstruction. The microinstruction contains a control word that specifies one or more microoperations for the data processor. The next address is computed in the next address generator also called microprogram sequencer.

Each word in control memory contains within it a microinstruction . The microinstruction specifies one or more microoperations for the system. A sequence of microinstructions constitutes a microprogram . Since alterations of the microprogram are not needed once the control unit is in operation, the control memory can be a read-only memory (ROM).

The content of the words in ROM are fixed and cannot be altered by simple programming since no writing capability is available in the ROM. ROM words are made permanent during the hardware production of the unit. The use of a microprogram involves placing all control variables in words of ROM for use by the control unit through successive read operations. The content of the word in ROM at a given address specifies a microinstruction. A more advanced development known as dynamic microprogramming permits a microprogram to be loaded initially from an auxiliary memory such as a magnetic disk. Control units that use dynamic microprogramming employ a writable control memory. This type of memory can be used for writing (to change the microprogram) but is used mostly for reading. A memory that is part of a control unit is referred to as a control memory. A computer that employs a microprogrammed control unit will have two separate memories: a main memory and a control memory. The main memory is available to the user for storing the programs. The contents of main memory may alter when the data are manipulated and every time that the program is changed. The user's program in main memory consists of machine instructions and data. In contrast, the control memory holds a fixed microprogram that cannot be altered by the occasional user. The microprogram consists of microinstructions that specify various internal control signals for execution of register microoperations. Each machine instruction initiates a series of microinstructions in control memory. These microinstructions generate the microoperations to fetch the instruction from main memory; to evaluate the effective address, to execute the operation specified by the instruction, and to return control to the fetch phase in order to repeat the cycle for the next instruction.

The control memory is assumed to be a ROM, within which all control information is permanently stored.

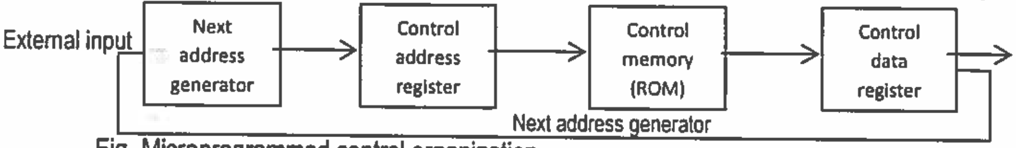


Fig. Microprogrammed control organization

10 (b) What do you mean by Complex Instruction Set Computer (CISC)? Discuss relative advantages and disadvantages of such instruction set design

CISC-1 M + Advantages and disadvantages of CISC-3M

Complex Instruction Set Computer (CISC) is a computer architecture in which single instructions can execute several low-level operations or are capable of multi-step operations or addressing modes within single instructions.

Example of CISC architectures are System/360, PDP-11,Motorola 6800.

CISC works by combining simple instructions into a single complex one, thereby optimizing the instructions per program and reducing the number of instructions that a particular program has.

Advantages of CISC:

- Microprogramming requires assembly language that is easier to implement.
- Instructions that manipulate operands in memory
- The CISC architecture reduces the amount of work that the compilers have to do because the instructions are already high-level
- Fewer instructions needed to implement a task

Disadvantages of CISC:

- The number of general-purpose registers that can be fitted into the processor is less because decoding instructions require more transistors
- As CPU does more work in a single instruction, the clock speed tends to be slightly slower than a RISC-based CPU
- The code requires several clock cycles to execute a single instruction despite having a minimal code size. This can decrease system efficiency.
- To simplify software hardware needs to be complex

11 (a) Explain the basic computer instruction formats

Computer instruction format-1 M + Types-Zero,One-address,Two-Address,Three-Address and RISC with examples-4M

Computer instruction format is defined as standard machine instruction format that can be directly decoded and executed by the CPU. It defines the layout of the instruction.

Instruction includes a set of operation codes and operands that manage with the operation codes. Instruction format contains fields including opcode, operands, and addressing mode.

Based on number of address, instructions are classified as zero-address, one-address, two-address, three-address and RISC.

Example : X = (A+B) * (C+D)

Zero Address Instruction

PUSH A	TOS←A
PUSH B	TOS←B
ADD	TOS← (A + B)
PUSH C	TOS←C
PUSH D	TOS←D
ADD	TOS←(C + D)
MUL	TOS← (C + D) X (A + B)
POP	M[X] ← TOS

One Address Instruction

LOAD A	AC←M[A]
ADD B	AC← AC+ M[B]
STORE T	M[T]←AC
LOAD C	AC←M[C]
ADD D	AC← AC+ M[D]
MUL T	AC← AC * M[T]
STORE X	M[X] ←AC

Two Address Instruction

MOV R1,A	R1←M[A]
ADD R1,B	R1←R1+M[B]
MOV R2,C	R1←M[C]
ADD R2,D	R2←R2+M[D]
MUL R1,R2	R1←R1*R2
MOV X,R1	M[X]←R1

Three Address Instruction

ADD R1, A, B	R1 ←M[A]+M[B]
ADD R2, C, D	R2 ←M[C]+M[D]
MUL X,R1,R2	M[X] ← R1 * R2

RISC Instruction

LOAD R1, A	R1←M[A]
LOAD R2, B	R2←M[B]
LOAD R3, C	R3←M[C]
LOAD R4, D	R4←M[D]
ADD R1,R1,R2	R1←R1+R2
ADD R3,R3,R2	R3←R3+R2
MUL R1,R1,R3	R1←R1*R3
STORE X,R1	M[X] ←R1

11 (b) What do you mean by addressing modes? Explain the following addressing modes

(i) Index addressing mode (ii) Immediate Addressing mode (iii) Relative Addressing mode (iv) Direct addressing mode

Addressing Mode-1 M + Index, immediate, relative, direct addressing modes-1.5 M for each addressing mode with examples

Addressing mode refers to the way in which the operand of an instruction is specified. The addressing mode specifies a rule for interpreting or modifying the address field of the instruction before the operand is actually executed.

Immediate addressing mode(Symbol #): In this mode data is present in the address field of the instruction itself.

Example : MOV AL,35H

Index addressing mode: In this mode data index register is added to the address part of the instruction to obtain the effective address.

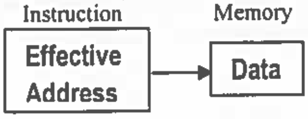
Example:LD ADR(X) AC←M[XR + ADR]

Relative addressing mode: In this mode program counter is added to the address field of the instruction to obtain the effective address. It is used to implement intra segment transfer of control.

ExampleLD $ADR AC←M[PC + ADR]

Direct addressing mode:

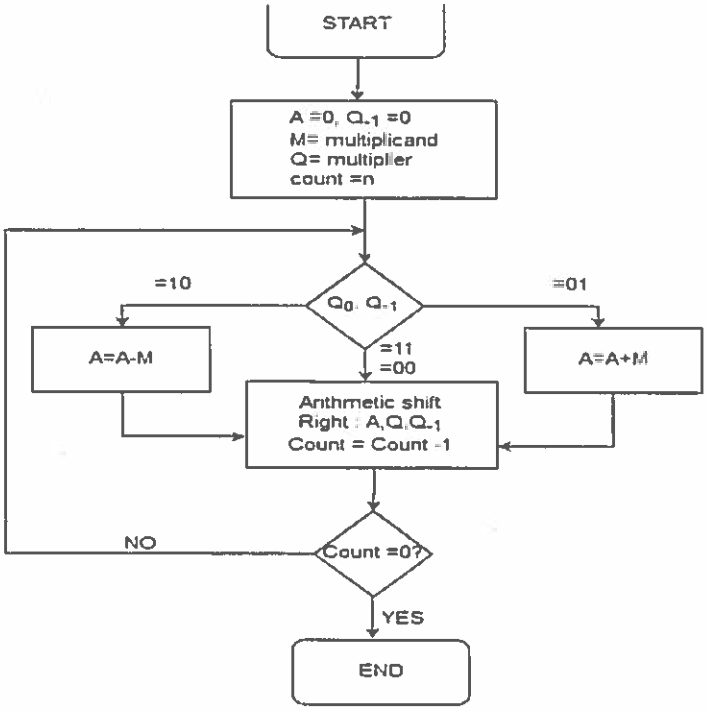
The effective address of the data is part of the instruction itself. Here only one memory reference is used to access the data as the address specifies which register or memory word contains the operand.



Example: ADD AL,[0123]. This instruction adds the contents of location 123 to AL and stores the result in AL

12 (a) Discuss Booth multiplication algorithm

Booth algorithm flowchart-4M+Example-2M



Here arithmetic shift right (ashr) is performed on A and multiplier Q including the appended Q_{-1}.

Example : Perform multiplication of 7 and 3

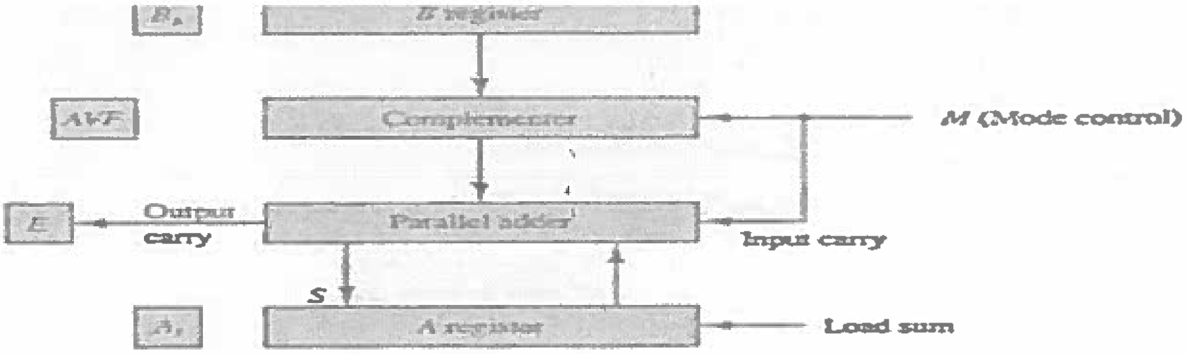
A	Q	Q_{-1}	M	SC	Remarks
0000	0011	0	0111	4	Initital values
1001	0011	0	0111	3	Subtract M { A←A – M }
1100	1001	1	0111		Perform Arithmetic shift right(ashr)
1110	0100	1	0111	2	Perform Arithmetic shift right(ashr)
0101	0100	1	0111	1	Addition { A←A + M }
0010	1010	0	0111		Perform Arithmetic shift right(ashr)
0001	0101	0	0111	0	Final product is available in A and Q

12 (b) Explain the functional units and their data flow in a hardware implementation that performs addition and subtraction of signed-magnitude numbers

Table for addition & subtraction of signed magnitude fixed point-2MHardware for signed magnitude addition and subtraction-4M

Operation	Add Magnitudes	Subtract Magnitudes		
		When $A > B$	When $A < B$	When $A = B$
$(+A) + (+B)$	$+(A + B)$			
$(+A) + (-B)$		$+(A - B)$	$-(B - A)$	$+(A - B)$
$(-A) + (+B)$		$-(A - B)$	$+(B - A)$	$+(A - B)$
$(-A) + (-B)$	$-(A + B)$			
$(+A) - (+B)$		$+(A - B)$	$-(B - A)$	$+(A - B)$
$(+A) - (-B)$	$+(A + B)$			
$(-A) - (+B)$	$-(A + B)$			
$(-A) - (-B)$		$-(A - B)$	$+(B - A)$	$+(A - B)$

Table-Addition and subtraction of signed-magnitude-numbers



13 (a) Draw a flowchart which explains multiplication of two signed magnitude fixed point numbers
Flowchart for multiplication of 2 signed magnitude fixed point numbers—6M

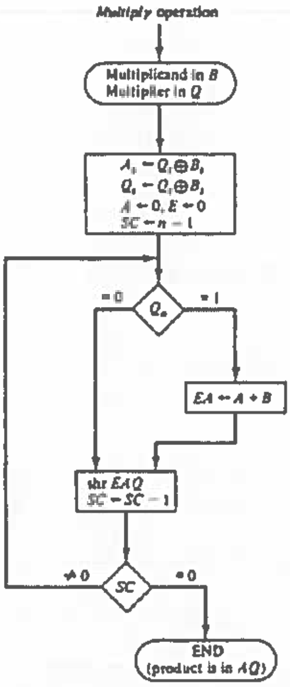


Figure Flowchart for multiply operation.

13 (b) Draw a flowchart for adding and subtracting two fixed point binary numbers where negative numbers in signed 1's complement representation
Flowchart for addition and subtraction of two fixed point binary numbers in signed 1's complement—6M

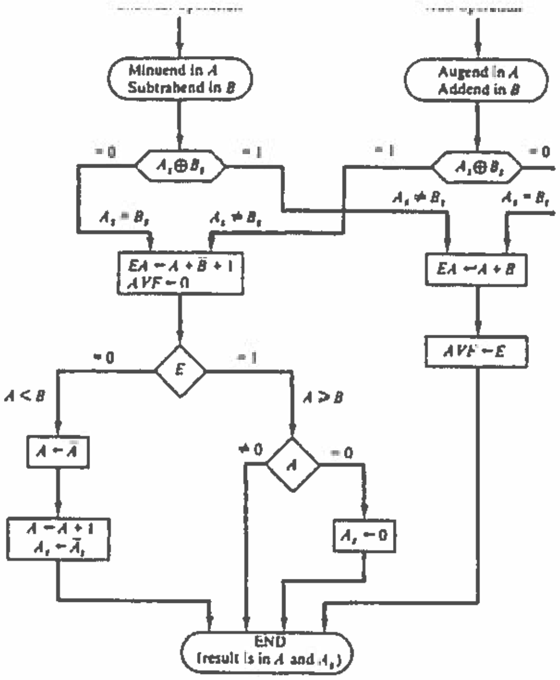
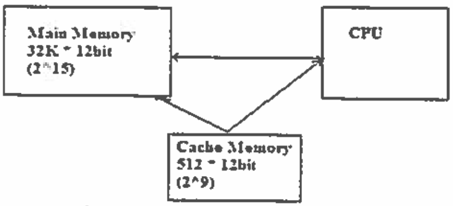


Figure Flowchart for add and subtract operations.

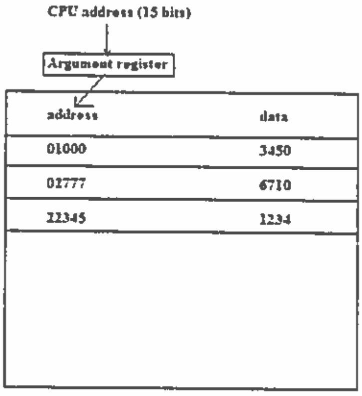
14 (a) What is the significance of cache memory and write about direct and associative mapping techniques
Significance of cache memory—1M + Direct Mapping —3M + Associative mapping—3M
Cache memory consumes less access time compared to main memory. Cache is faster than main memory and hence CPU first refers to any data if available in cache, if available it is called a hit and improves the performance of the system.
The transformation of data from main memory to cache memory is referred to as a mapping process. There are three types of mapping:

i) Associative mapping
ii) Direct mapping
iii) Set-associative mapping

To help understand the mapping procedure,we have the following example:



Associative mapping: The fastest and most flexible cache organization uses an associative memory. The associative memory stores both the address and data of the memory word. This permits any location in cache to store any word from main memory
The address value of 15 bits is shown as a five-digit octal number and its corresponding 12- bit word is shown as a four-digit octal number.



If the address is found, the corresponding 12- bits data is read and sent to the CPU. If not, the main memory is accessed for the word. If the cache is full, an address-data pair must be displaced to make room for a pair that is needed and not presently in the cache.

Direct Mapping

Associative memory is expensive compared to RAM. In general case, there are 2^k words in cache memory and 2^n words in main memory (in our case, k=9, n=15). The n bit memory address is divided into twofields: k-bits for the index and n-k bits for the tag field.

Each word in cache consists of the data word and its associated tag. When a new word is first brought into the cache, the tags are stored alongside the data bits. When the CPU generates a memory request, the index field is used for the address to access the cache.

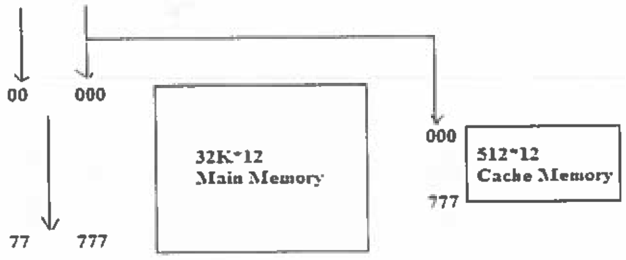
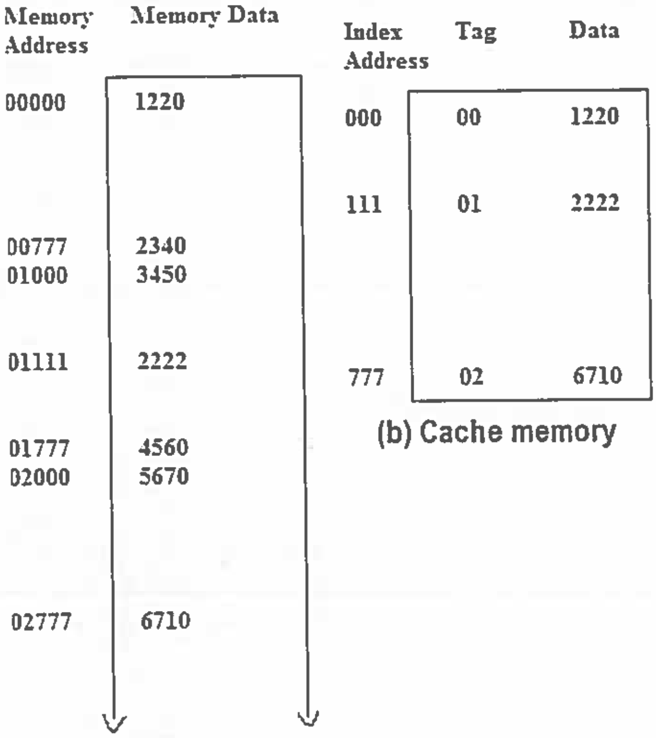


Fig. Addressing relationships between main and cache memories



Memory Address	Memory Data
00000	1220
00777	2340
01000	3450
01111	2222
01777	4560
02000	5670
02777	6710

(a) Main memory

Index Address	Tag	Data
000	00	1220
111	01	2222
777	02	6710

(b) Cache memory

Fig. Direct mapping cache organization

The tag field of the CPU address is compared with the tag in the word read from the cache. If the two tags match, then it's a hit and the desired data word is in cache. Otherwise it's a miss and required word is read from main memory. It is then stored in the cache together with the new tag, replacing the previous value.

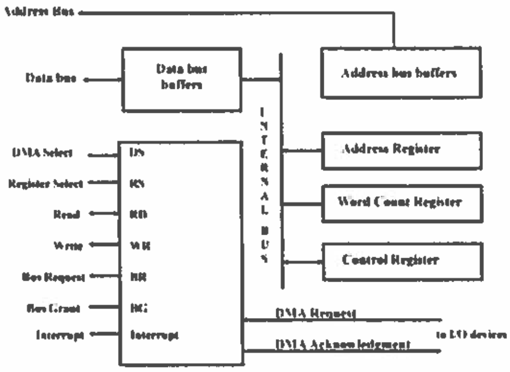
Consider the numerical example shown above. The word at address zero is presently stored in the cache(index=000, tag=00, data=1220). Let us assume that CPU wants to access the word at address 01000. The index address is 000, so it is used to access the cache. The two tags are then compared. The cache tag is 00 but the address tag is 01, which does not produce a match. Hence, the main memory is accessed and the data word 3450 is transferred to the CPU. The cache word at index address 000 is then replaced with a tag of 01 and data of 3450.

14 (b) Distinguish between synchronous and asynchronous data transfer
Synchronous data transfer—1 M + Asynchronous data transfer—1 M+ Differences between them—3 M
In synchronous data transfer the sending and receiving units are enabled with the same clock signal

Synchronous data transfer	Asynchronous data transfer
The master performs a sequence of instructions for data transfer in a predefined order	None of the actions are synchronized with a common clock, hence no predefined order
As there is common clock there is no need of control signal between source and destination while transferring data	The asynchronous data transfer between the two units require control signals be transmitted between the communicating units to indicate when to send data
The master does not expect any acknowledgement when data is sent by the master to the slave	In handshaking method there is possibility of knowing whether sent data is received or not at the receiver end via acknowledgement

15 (a) With the help of a block diagram, explain the concept of DMA controller
DMA controller—1M + Block diagram of DMA controller with explanation—5M

DMA Controller:
The DMA controller needs the usual circuits of an interface to communicate with the CPU and I/O device. The DMA controller has three registers :Address Register, Word Count Register and Control Register
Address Register :- Address Register contains an address to specify the desired location in memory.
Word Count Register(WC) :- WC holds the number of words to be transferred. The register is incremented/decremented by one after each word transfer and internally tested for zero.
Control Register :- Control Register specifies the mode of transfer
The unit communicates with the CPU via the data bus and control lines. The registers in the DMA are selected by the CPU through the address bus by enabling the DS (DMA select) and RS (Register select) inputs. The RD (read) and WR (write) inputs are bidirectional.



Block Diagram of DMA Controller

When the BG (Bus Grant) input is 0, the CPU can communicate with the DMA registers through the data bus to read from or write to the DMA registers. When BG =1, the DMA can communicate directly with the memory by specifying an address in the address bus and activating the RD or WR control.

15 (b) What is virtual memory? With a neat block diagram explain the virtual memory address translation

Virtual memory—1M + Virtual memory address translation with neat diagram—5M
A computer can address more memory than the amount physically installed on the system. This extra memory is actually called virtual memory and it is a section of a hard disk that's set up to emulate the computer's RAM.

A virtual memory system provides a mechanism for translating program-generated addresses into correct main memory locations. This is done dynamically, while programs are being executed in the CPU. The translation or mapping is handled automatically by the hardware by means of a mapping table.

An address used by a programmer will be called a virtual address, and the set of such addresses the address space.

An address in main memory is called a location or physical address. The set of such locations is called the memory space.

For example, consider a computer with a main-memory capacity of 32K (2^5 X 2^{10}) words .15 bits are needed to specify a physical address in memory . Suppose that the computer has available auxiliary memory for storing 2^{20} = 1024K words. Thus auxiliary memory has a capacity for storing information equivalent to the capacity of 32 main memories. For this example address space N= 1024K and memory space M = 32K.

In a multiprogram computer system, programs and data are transferred to and from auxiliary memory and main memory based on demands imposed by the CPU. Suppose that program 1 is currently being executed in the CPU.

In a virtual memory system the address field of the instruction code has a sufficient number of bits to specify all virtual addresses. In our example, the address field of an instruction code will consist of 20 bits but physical memory addresses must be specified with only 15 bits. Thus CPU will reference instructions and data with a 20-bit address, but the information at this address must be taken from physical memory because access to auxiliary storage for individual words will be prohibitively long.

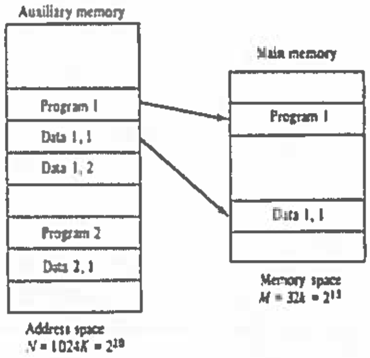


Figure Relation between address and memory space in a virtual memory system.

A table is needed to map a virtual address of 20 bits to a physical address of 15 bits.

The mapping table may be stored in two ways:

1) Stored in a separate memory (additional memory) -Here an additional memory unit is required as well as one extra memory access time

2) Main memory –Here the table takes space from main memory and two accesses to memory are required with the program running at half speed.

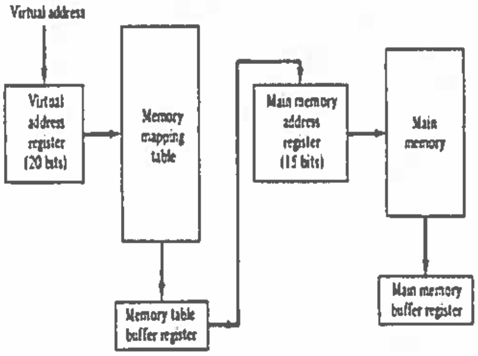
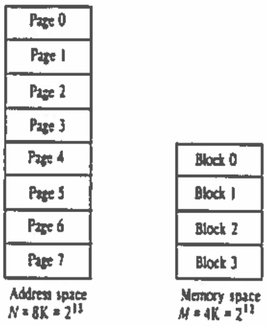


Figure Memory table for mapping a virtual address.

The physical memory is broken down into groups of equal size called blocks, which may range from 64 to 4096 words each. The term "page frame" is sometimes used to denote a block. The term page refers to groups of address space of the same size. Consider a computer with an address space of 8K and a memory space of 4K. If we split each into groups of 1K words we obtain eight pages and four blocks as shown in Fig.

At any given time, up to four pages of address space may reside in main memory in any one of the four blocks.



Address space and memory space split into groups of 1K words.

Address mapping using pages and tables

The table implementation of the address mapping is simplified if the information in the address space. And the memory space is each divided into groups of fixed size.

The physical memory is broken down into groups of equal size called blocks, which may range from 64 to 4096 words each.

The term page refers to groups of address space of the same size.

Also, consider a computer with an address space of 8K and a memory space of 4K.

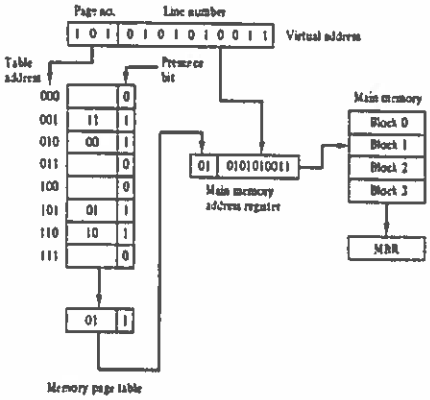


Figure Memory table in a paged system.

Consider a computer with an address space of 8K and a memory space of 4K.

If we split each into groups of 1K words we obtain eight pages and four blocks as shown in the figure.

If we split each into groups of 1K words we obtain eight pages and four blocks as shown in the figure.

At any given time, up to four pages of address space may reside in main memory in any one of the four blocks.

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2022 - 2023
Course Code	20EE305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Power Generation and Transmission**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any two disadvantages of nuclear power plant.	20EE305.1	L1
2	Define the diversity factor.	20EE305.2	L1
3	Define GMR and GMD.	20EE305.3	L1
4	Recall surge impedance loading (SIL).	20EE305.4	L1
5	Mention any two methods of improving string efficiency.	20EE305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Explain the construction and working principle of thermal power plant.	12M	20EE305.1	L2
	OR			
7	Explain the layout, classification, and operation of hydro power plant.	12M	20EE305.1	L2
8	A generating station has the following daily load cycle: (table below) 70MW 1200 MWh Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor. 50MW 71.4	12M	20EE305.2	L3
	OR			
9	Explain the types of Tariff methods.	12M	20EE305.2	L2
10	Derive the equation for inductance of a three phase over head line.	12M	20EE305.3	L3
	OR			
11	Derive the equation for capacitance of a two-wire over head line.	12M	20EE305.3	L3
12	Classify the types of transmission lines with model representations.	12M	20EE305.4	L2
	OR			
13 (a)	Derive the expressions for the Performance of long transmission lines using rigorous method with relevant equations.	6M	20EE305.4	L3
13 (b)	Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.	6M	20EE305.4	L3
14	Explain the different methods used to improve the string efficiency of insulators with necessary diagrams.	12M	20EE305.5	L2

Question 8 load cycle:

Time (Hrs)	0-6	6-10	10-12	12-16	16 -20	20 -24
Load (MW)	40	50	60	50	70	40

	OR			
15 (a)	Derive an expression for sag of a line supported between two supports of the same tower height.	8M	20EE305.5	L3
15 (b)	A 132 kV transmission line has the following data: Wt. of conductor = 680 kg/km; Length of span = 260 m Ultimate strength = 3100 kg; Safety factor = 2. Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 m.	4M	20EE305.5	L3

Height = 13.7M

Tension , T = 1550

d = 3.7 m

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ANSWER KEY AND SCHEME OF EVALUATION

Part - A:

Q1: Two ~~ad~~ disadvantages — 2M. (1M for each).

Q2: Definition — 1M
Formula — 1M.

Q3: GMR – 1M
GMD – 1M

04: SIL Definition – 2M.

05: Two methods – 1M for each.

Part - B

Q6: Figure : 4M
Components List : 4M
Explanation : 4M.

Q7: Figure : 4M
Component List : 4M
Explanation : 4M.

Q8: (i) 2M ; (ii) 3M ; (iii) 2M ; (iv) 2M
Curve : 3M.

Q9: 6 Types 2M for each types.

Q10: Basic explanation : 4M.

Unsymmmetrical Derivation : 4M.

Transposition Derivation : 4M.

Q 11: Explanation of Law : 4M.

Derivation : 8M.

Q 12: Short T/m Line : 2M

Medium T/m Line : 6M.

Long T/m Line : 2M.
DC T/m line : 2M

Q 13a : Derivation : 6M.

b : Figures : 3M.
Derivation : 3M.

Q14 : Each Method : 3M.

Q 15 a : Figure : 4M
Derivation : 4M.

b : Complete Solution : 4M.

1. List any two disadvantages of nuclear power plant.
   Any two of the below (2m)
   a) Strong pressure vessel is required due to the use of high-pressure water system
   b) Formation of low temperature steam
   c) Use of expensive cladding material for prevention of corrosion
   d) 4.High losses from heat exchanger
   e) 5.High power consumption by auxiliaries
   f) 6.Requires more elaborate safety devices

2. Define the diversity factor.
   It shows the diversity of load connected to a power station (2m)
   Diversity factor = Sum of individual maximum demands / Maximum demand on power system

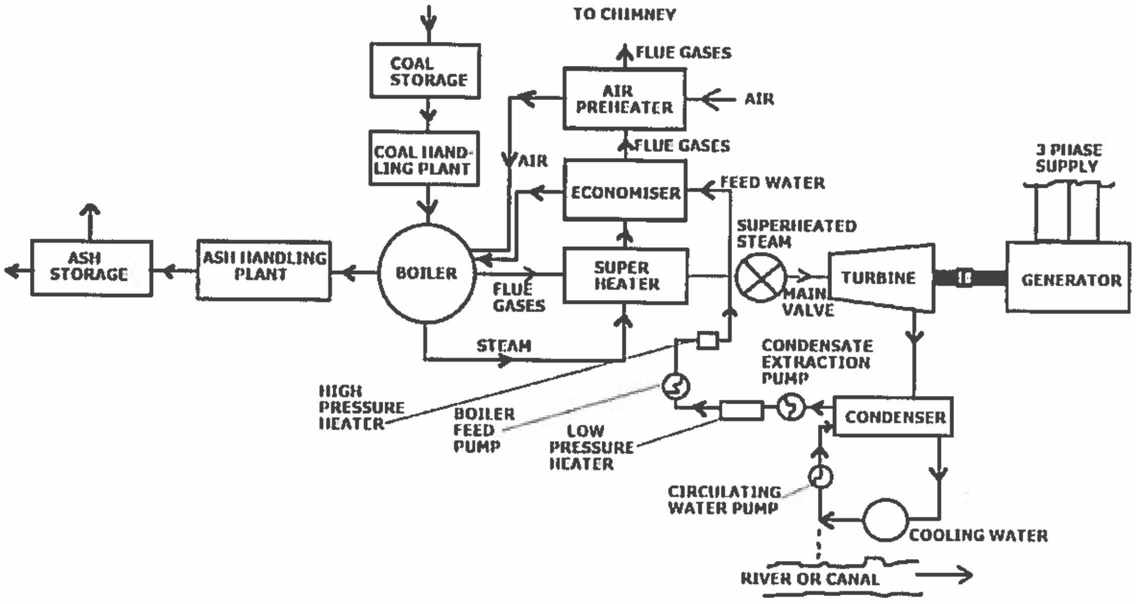
3. Define GMR and GMD.
   GMR stands for Geometrical Mean Radius. It is also called the self GMD (Geometrical Mean Distance)
   GMR = 0.7788R
   Where R is the radius of the conductor (1m)

   GMD stands for Geometrical Mean Distance GMD represents the geometrical mean distance from one conductor to the other. GMD for a different arrangement of conductors has different values. (1m)

4. Recall surge impedance loading (SIL).
   SIL is defined as the maximum load (at unity power factor) that can be delivered by the transmission line when the loads terminate with a value equal to surge impedance (Zs) of the line. (2m)

5. Mention any two methods of improving string efficiency.
   Any of the following (2m)
   By Using Insulators with Larger Discs or by Providing Each Insulator Unit with a Metal Cap
   By Using Longer Cross-Arms
   By Capacitance Grading
   By Static Shielding

6. Explain the construction and working principle of thermal power plant.



A Thermal Power plant converts the heat energy of coal into electrical energy. Coal is burnt in boiler which converts water into steam.

The expansion of steam in turbine produces mechanical power which drives the Alternator coupled to the turbine. Thermal power plants contribute maximum to the generation of power for any country.

In thermal generating stations coal, oil, natural gas etc. are employed as primary courses of energy.

Main Components

- Coal handling plant
- Pulverizing plant
- Boiler
- Turbine
- Condenser
- Cooling towers and ponds
- Feed water heater
- Economizer
- Air preheater

Coal Handling Plant

- Coal is transported to power station by rail or road and stored in coal storage plant and then pulverized.
- The function of coal handling plant is automatic feeding of coal to the boiler furnace.
- A thermal power plant burns enormous amounts of coal.
- A 200MW plant may require around 2000 tons of coal daily.

Pulverizing Plant

- In modern thermal power plant, coal is pulverized i.e. ground to dust like size and carried to the furnace in a stream of hot air. Pulverizing is a means of exposing a large surface area tot the action of oxygen and consequently helping combustion.
- Pulverizing process consists 3 stages classified as:
  1.Feeding
  2.Drying.
  3.Grinding

Boiler

- The function of boiler is to generate steam at desired pressure and temperature by transferring heat produced by burning of fuel in a furnace to change water into steam.

Turbine

- In thermal power plants generally 3 turbines are used to increase the efficiency.
- High pressure turbine
- Intermediate pressure turbine
- Low pressure turbine

Condenser

- The surface condenser is a shell and tube heat exchanger where cooling water flows through tubes and exhaust steam fed into the shell surrounds the tubes, as a result, steam condense outside the tubes.

Cooling Towers and Ponds

- A condenser needs huge quantity of water to condense the steam.
- Most plants use cooled cooling system where warm water coming from condenser is cooled and reused.
- Cooling tower is a steel or concrete hyperbolic structure with the height of 150m.

Feed water heater

- Feed water heating improves overall plant efficiency
- Thermal stresses due to cold water entering the boiler drum are avoided.
- Quality of steam produced by the boiler is increased.

Economizer

- Flue gases coming out of the boiler carry lot of heat. An economizer extracts a part of this heat from flue gases and uses it for heating fee water.

- Saving coal consumption and higher boiler efficiency.

Air Preheater

- The function of air preheaters is to preheat the air before entering to the furnace by utilizing some of the energy left in the flue gases before exhausting them to the atmosphere.
- After flue gases leave economizer, some further heat can be extracted from them and used to heat incoming heat. Cooling of flue gases by 20-degree centigrade increases the plant efficiency by 1%.

Ash Handling plant

- The ash from the boiler is collected in two forms
- Bottom ash (Slurry): It's a waste which is dumped into ash pond
- Fly ash: Fly ash is separated from flue gases in ESP.

Water Handling Plant

- Water in a power plant is used for
- Production of steam - for rotating turbine
- Cooling purpose - For cooling of various equipment.
- Water is recycled and used for various purpose:
  Raw water - For cooling purposes - Steam - Condenser - Raw water
- About 4 cubic meter water is lost/day/MW.

Electrostatic precipitator (ESP)

- An ESP electrically charges the ash particles and imparts a strong electric field in the flue gas to collect and remove them. ESP is comprised of a series of parallel, vertical metallic plates (collecting electrodes) forming lanes through which the flu gases pass.

7. Explain the layout, classification, and operation of hydro power plant.

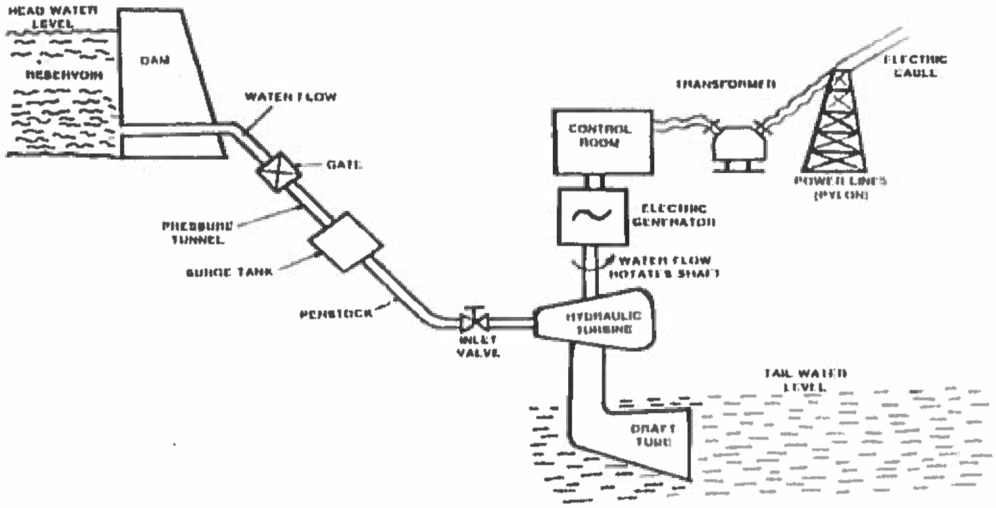


Fig.Layout of Hydro electric Power plant

(4m)

In hydroelectric power station potential and kinetic energy of stored water is converted into electrical energy.

Working Principle

- Hydroelectric power is the power obtained from the energy falling water whereas hydroelectric power plant is the power utilizing the potential energy of water at a high level for the generation of the electrical energy.

Main Components

- Reservoir
- Dam
- Trace rack
- Turbine
- Forebay

- Surge tank
- Penstock
- Spillway
- Turbine
- Powerhouse (4m)

Dam

- Develops a reservoir to store water
- Builds up head for power generation

Spillway

- To safeguard the dam when water level in the reservoir rises

Intake

- Contains trash racks to filter out debris which may damage the turbine

Forebay

- The forebay is used as "regulating reservoir" storing water temporarily during light load and providing the same for initial increases on account of increasing load during which water in canal is being accelerated. A forebay is an enlarged body of water just above the intake which is used to store water temporarily to meet the hourly load fluctuations. A forebay is not required if plant is located just at the base of the dam but, if the plants are situated away from the storage reservoir, a forebay is must.

Surge tank

- Surge tank is a small reservoir in which the water level rises or falls to reduce the pressure swings so that they are not transmitted to the penstock.
- When the load demand is reduced on the power station then, it causes rise in water level in the surge tank which produces a retarding head and reduces the velocity of water in the penstock and hence avoiding the undesirable phenomenon called "water hammer"
- When the load on the plant is increased, the governor causes the turbine to open the gates in order to allow more water to flow through the penstock to supply the increased load and there is a tendency to cause a vacuum or a negative pressure in the penstock.

Penstock

- Penstock is a closed conduit which connects the forebay or surge tank to the scroll case of the turbine. In case of high head plants, a single penstock is provided.

Valves and Gates

- Gates are used in low head plants at the entrance to the turbine casing to shut-off the flow and provide for unwatering the turbine for inspection and repairs. Valves are used at the entrance to the turbine casing if a long or medium length penstocks is used in the hydro power plant.

Trash Racks

- Trash racks are used to prevent the ingress of floating and other material to the turbine. These are built up from long, flat bars set vertically or nearly so and spaced in accordance with the minimum width of water passage through the turbine.

Tail race

- After the useful work is done by water, it is discharged to the tail race.

Draft tube

- It is an airtight pipe of suitable diameter attached to the runner outlet and conducting water down from the wheel and discharging it under the surface of the water in the tail race. With the help of draft tube operating head on the turbine is increased resulting in increase in output and efficiency

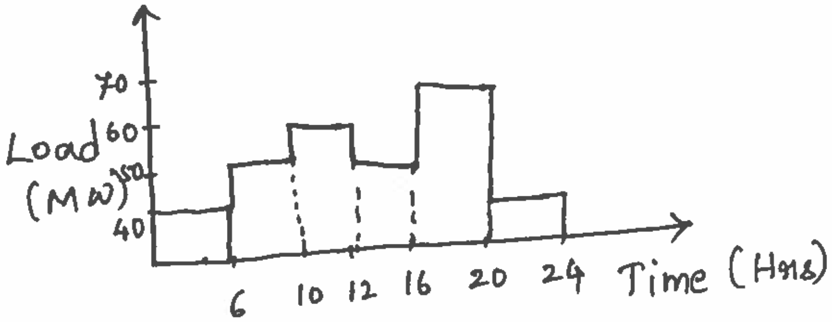
Prime Movers or water turbines

- In hydroelectric plants water turbines serve the purpose of prime mover which converts the kinetic energy of water into mechanical energy which is further utilized to drive the alternators generating electrical energy. (4m)

8. A generating station has the following daily load cycle:

Time (Hrs)	0-6	6-10	10-12	12-16	16 -20	20 -24
Load (MW)	40	50	60	50	70	40

Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor.



(i) Maximum Demand = 70MW.

(ii) Units Generated per day = $(6\times40)+(4\times50)+(2\times60)$
$+(4\times50)+(4\times70)+(4\times40)$

= 1200 MWh.

(iii) Average Load = $\dfrac{\text{Units generated/day}}{\text{Hrs in a day}}$

$= \dfrac{1200\ \text{MWh}}{24} = \dfrac{1200\times10^6}{24}$

$= 50\times10^6 = 50\text{MW}.$

(iv) Load Factor = $\dfrac{\text{Average Demand}\times100}{\text{Maximum Demand}} = \dfrac{50\text{MW}}{70\text{MW}}\times100$

$= 0.714\times100$

$= 71.4.$

9. Explain the types of Tariff methods.

Simple Tariff: When there is a fixed rate per unit of energy consumed, it is the Simple Tariff. In this type of tariff, the price charged per unit is constant. It does not vary with increase or decrease with the number of Units consumed.

Flat Rate Tariff: When different types of Consumers are charged at different uniform per unit rates, it is the Flat Rate Tariff. The rate for each type of consumer is arrived at by taking its load factor, diversity factor into consideration. The Bill will be Total units Consumed x Rate/Unit.

Block Rate Tariff: When a given block of energy is charged at a specific rate and the succeeding blocks of energy are charged progressively at reduced rates. Then the Tariff is called the Block Rate Tariff.

If the number of units generated increases, then the cost of generation per-unit-automatically decreases. For the first 30 units may be charged at the rate of 60paisa per unit, the next 25 units at the rate of 55paise per unit and the remaining additional units may be charged at the rate of 30 paise per unit. This type of tariff is being majorly used for residential and small commercial consumers.

Two Part Tariff: When the rate of electric energy is charged based on maximum demand of the consumer and the units consumed, it is called the Two-part Tariff.

In two-part tariff, the total charge to be made from the consumer is spilt into two components, fixed charges and running charges. The fixed charges depend upon the maximum demand of the consumer, while the running charges depend upon the no. of units consumed by the consumer.

Total Charges = Rs. (B kW+ C kWh)

B = Charges per kW of maximum demand

C = Charges per kWh of energy consumed

This type of Tariff is mostly applicable to Industrial Consumers.

Maximum Demand Tariff: It is similar to Two Part Tariff with the only difference that the maximum demand is actually measured by installing maximum demand meter in the premises of the consumer.

This type of tariff is mostly applied to big consumer, as a separate maximum demand meter is required.

Three-part tariff: When the total charge to be made from the consumer split into three parts, fixed charges, Semi fixed charges and running Charges.

Total Charge = Rs. (A + B kW + C kWh)

A = Fixed Charges during each billing period

B = Charge per kW of Maximum demand

C = Charge per kWh of energy consumed

Power Factor Tariff:

The Tariff in which Power factor of the Consumer's load is taken into consideration is known as Power Factor Tariff.

The following are the important types of Power Factor tariff

(i) kVA maximum demand tariff: It is a modified form of two-part tariff. The fixed charges are made on the basis of maximum demand in kVA and not in kW. As kVA is inversely proportional to power factor, a consumer having low power factor has to contribute more towards the fixed charge.

(ii) Sliding Scale Tariff: This is also known as average PF tariff. In this case an average power factor say 0.8 lagging is taken as reference. If the pf of the consumer falls below this factor, suitable additional charges are made. If the PF is above the reference, a discount is allowed to the consumer.

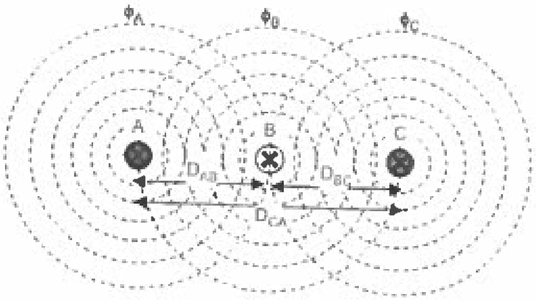
(iii) kW and kVAR Tariff: In this type both active power (kW) and reactive power (kVAR) supplied are charged separately. A consumer having low PF will draw more reactive Power and hence shall have to pay more charges.

10. Derive the equation for inductance of a three-phase overhead line.

Inductance in Three Phase Transmission Line:

In the three-phase transmission line, three conductors are parallel to each other. The direction of the current is same through each of the conductors.

Let us consider conductor A produces magnetic flux φ_A, Conductor B produces magnetic flux φ_B, And conductor C produces magnetic flux φ_C. When they carry the current of the same magnitude "I", they are in flux linkage with each other.

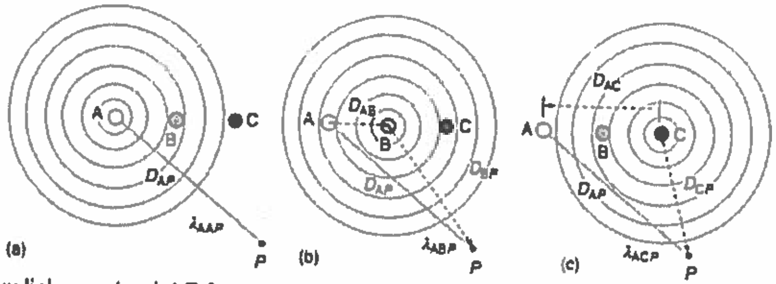


Now, let us consider a point P near three conductors. So, flux linkage at point P due to current through conductor A is,

$$\lambda_{AP} = \lambda_{AAP} + \lambda_{ABP} + \lambda_{ACP}$$

Flux linkage at point P for conductor A due to current through conductor A =

$$\lambda_{AAP} = \frac{\mu_0}{2\pi} I_A \times ln\left(\frac{D_{AP}}{GMR_A}\right) \quad Wb/m$$



Flux linkage at point P for conductor A due to current through conductor B =

$$\lambda_{ABP} = \frac{\mu_0}{2\pi} I_B \times ln\left(\frac{D_{BP}}{D_{AB}}\right) \quad Wb/m$$

Flux linkage at point P for conductor A due to current through conductor C =

$$\lambda_{ACP} = \frac{\mu_0}{2\pi} I_C \times ln\left(\frac{D_{CP}}{D_{AC}}\right) \quad Wb/m$$

Therefore, flux linkage at point P for conductor A,

$$\Rightarrow \lambda_{AP} = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right] + \frac{\mu_0}{2\pi}\left[I_A \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})\right] \quad Wb/m$$

(2m)

As, $D_{AP} = D_{BP} = D_{CP}$ and $I_A + I_B + I_C = 0$ in balanced system, then we can write that $I_A = -I_B - I_C$

$$\therefore \frac{\mu_0}{2\pi}[I_A \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})]$$

$$= \frac{\mu_0}{2\pi}[-I_B \times ln(D_{AP}) - I_C \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})] = 0$$

$$\Rightarrow \lambda_{AP} = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right] + 0 = \lambda_A(say)$$

$$So,\ \lambda_A = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right]$$

$$Similarly,\ \lambda_B = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{D_{BA}}\right) + I_B \times ln\left(\frac{1}{GMR_B}\right) + I_C \times ln\left(\frac{1}{D_{BC}}\right)\right]$$

$$and,\ \lambda_C = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{D_{CA}}\right) + I_B \times ln\left(\frac{1}{D_{CB}}\right) + I_C \times ln\left(\frac{1}{GMR_C}\right)\right]$$ (2m)

For a balanced system,

$$D_{AB} = D_{BC} = D_{CA} = D$$

$$I_A + I_B + I_C = 0$$

In balanced system, then we can write that, $I_A = -I_B - I_C$

$$\lambda_A = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D}\right) + I_C \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + (I_B + I_C) \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + (-I_A) \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi} I_A \times ln\left(\frac{D}{GMR_A}\right) \quad Wb/m$$

$$\lambda_B = \frac{\mu_0}{2\pi} I_B \times ln\left(\frac{D}{GMR_B}\right) \quad Wb/m$$

$$\lambda_C = \frac{\mu_0}{2\pi} I_C \times ln\left(\frac{D}{GMR_C}\right) \quad Wb/m$$

(2m)

$$So,\ inductance\ per\ metre\ per\ phase,\ L_{phase} = \frac{\mu_0}{2\pi} \times ln\left(\frac{D}{GMR_{phase}}\right) \quad H/m$$

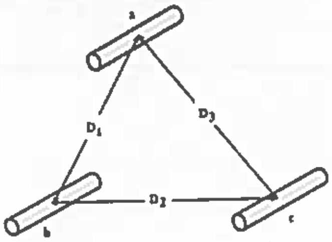
Consider a 3-phase overhead transmission line with phase conductors a, b, and c of radius r each spaced unsymmetrically such that the distance between the three conductors is D_1, D_2, and D_3 as shown in the figure below. Let I_a, I_b, and I_c be the currents flowing in the conductors a, b, and c respectively.

Assume that the whole system is balanced which leads to the flow of equal currents in the conductors i.e., $I_a = I_b = I_c = I$(say). The currents I_a, I_b, and I_c are displaced at 120° apart from each other. If I_a is taken as the reference phasor then the currents are given by,

$$I_a = I(1 - j0)$$

$$I_b = I(-0.5 - j0.866)\ and$$

$$I_c = I(-0.5 + j0.866)$$



Unsymmetrically Spaced 3-Phase Conductors

We know that the flux linkage of any conductor, in a group of conductors, is due to its own current and currents in the other conductors. Therefore, flux linkage of the conductor 'a' is due to its own current and currents in the conductor's 'b' and 'c', and it is given by,

$$\lambda_a = 2 \times 10^{-7}\left[I_a \ln\frac{1}{r'} + I_b \ln\frac{1}{D_1} + I_c \ln\frac{1}{D_3}\right]$$

Similarly, flux linkages of conductors 'b' and 'c' is given by,

$$\lambda_b = 2 \times 10^{-7}\left[I_b \ln\frac{1}{r'} + I_a \ln\frac{1}{D_1} + I_c \ln\frac{1}{D_2}\right]$$

$$\lambda_c = 2 \times 10^{-7}\left[I_c \ln\frac{1}{r'} + I_a \ln\frac{1}{D_3} + I_b \ln\frac{1}{D_2}\right]$$

(2m)

Where, r' = Geometric mean radius (GMR) of conductor = 0.07788 × r. Substituting the values of I_a, I_b, and I_c in the above equation we get,

$$\lambda_a = 2 \times 10^{-7} \left[I \ln\frac{1}{r'} + I(-0.5 - j0.866) \ln\frac{1}{D_1} + I(-0.5 + j0.866) \ln\frac{1}{D_3} \right]$$

$$\lambda_a = 2 \times 10^{-7}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1} + \ln\sqrt{D_3} + j0.866 \ln D_1 + j0.866 \ln\frac{1}{D_3} \right]$$

$$\lambda_a = 2 \times 10^{-7}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\frac{\sqrt{3}}{2} \ln D_1 + j\frac{\sqrt{3}}{2} \ln\frac{1}{D_3} \right]$$

$$\lambda_a = 2 \times 10^{-7}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\sqrt{3} \ln \sqrt{\left(\frac{D_1}{D_3}\right)} \right]$$

We know that the inductance L_a is given by,

$$L_a = \frac{\lambda_a}{I_a}$$

$$L_a = \frac{2 \times 10^{-7}}{I}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\sqrt{3} \ln \sqrt{\left(\frac{D_1}{D_3}\right)} \right]$$

$$\therefore L_a = 2 \times 10^{-7} \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\sqrt{3} \ln \sqrt{\left(\frac{D_1}{D_3}\right)} \right] H/m$$

Similarly, the inductance due to the conductor's 'b' and 'c' can be calculated and they are given by,

$$L_b = 2 \times 10^{-7} \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_2} + j\sqrt{3} \ln \sqrt{\left(\frac{D_2}{D_1}\right)} \right] H/m$$

$$L_c = 2 \times 10^{-7} \left[\ln\frac{1}{r'} + \ln\sqrt{D_2 D_3} + j\sqrt{3} \ln \sqrt{\left(\frac{D_3}{D_2}\right)} \right] H/m$$

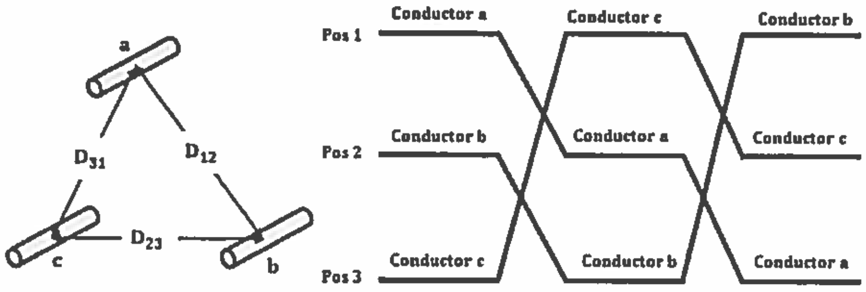
When conductors are unsymmetrically spaced in a 3-phase line, the flux linkage and inductance per phase are not identical, and knowing the inductance of each phase becomes complicated and it results in an unbalanced circuit.

The equilibrium in the circuit can be retained by shifting the places of the conductor at every period such that the conductors take the initial place of every conductor at the same distances.

Such an arrangement of the conductor's position obtained by shifting their places is called transposition. By transposition, we can obtain almost the same inductance between the conductors. Let us see the expression for inductance when the 3-phase line is transposed.

Inductance of Unsymmetrically Spaced 3-Phase Transmission Line When Transposed: (2m)

The simple configuration of a 3-phase conductor is shown in the figure. As the conductors are transposed, so their position in the transposition cycle would be as shown in the figure below.



Unsymmetrically Spaced 3-Phase Conductors with Transposition Cycle

When the conductors are connected in parallel, it results in low inductance keeping the distance between phases as small as possible. For deriving the inductance per phase of 3-phase conductors placed unsymmetrically, we need to determine the flux linkage of each conductor for every position it occupies in a transposition cycle. After that, we can determine the average flux linkages. So, the flux linkage of phase 'a' in position 1 is given as,

$$\lambda_{a1} = 2 \times 10^{-7}\left[I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{12}} + I_c \ln\frac{1}{D_{31}}\right]$$

When 'a' is in position 2, 'b' in position 3 and 'c' in position 1,

$$\lambda_{a3} = 2 \times 10^{-7}\left[I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{31}} + I_c \ln\frac{1}{D_{23}}\right]$$

The average value of flux linkage of a single-phase 'a' is,

$$\lambda_a = \frac{\lambda_{a1} + \lambda_{a2} + \lambda_{a3}}{3}$$

$$\lambda_a = \frac{2 \times 10^{-7}}{3}\left[3I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{12}D_{23}D_{31}} + I_c \ln\frac{1}{D_{12}D_{23}D_{31}}\right]$$

We know that

$$I_a + I_b + I_c = 0$$

$$I_b + I_c = -I_a$$

$$\therefore \lambda_a = \frac{2 \times 10^{-7}}{3}\left[3I_a \ln\frac{1}{D_s} - I_a \ln\frac{1}{D_{12}D_{23}D_{31}}\right]$$

$$\lambda_a = 2 \times 10^{-7} I_a \ln\sqrt{\frac{\sqrt[3]{D_{12}D_{23}D_{31}}}{D_s}}$$

Therefore, the average inductance per phase is,

$$L_a = 2 \times 10^{-7} \ln\frac{D_{eq}}{D_s}\ H/m$$

$$Where, D_{eq} = \sqrt[3]{D_{12}D_{23}D_{31}}$$

Where,

D_{eq} = Geometric mean of three distances of unsymmetrical line

D_s = GMR (geometric mean radius) of the conductor.

The above expression is the inductance per phase of a 3-phase transmission line with unsymmetrical spacing but lines are transposed. Nowadays the transposition of conductors is made at switching stations to balance inductance.

If the conductors are equispaced,

D1 = D2 = D3 = D

$$L = 2\,x\,10^{-7} \ln\frac{\sqrt[3]{d^3}}{r'} \Rightarrow L = 2\,x\,10^{-7} \ln\frac{d}{r'}$$ (2m)

11. Derive the equation for capacitance of a two-wire overhead line.

Electric Field Intensity due infinite line charge:

Consider a long wire having q coulomb/m as shown in figure

Using Gauss's law, Field Intensity (E) at a point P, which is r metre from the conductor, can be calculated as

$$\oint \overline{D} \cdot \overline{ds} = Q$$

The Flux density at point P, considering the cylindrical shell of radius r and length l can be calculated using Gauss's Law

$$D \cdot 2\Pi r \cdot l = ql \Rightarrow D = \frac{q}{2\pi r}\ coulomb/m^2$$

And Electric Field intensity E is given by

$$E = \frac{D}{\varepsilon_0}$$

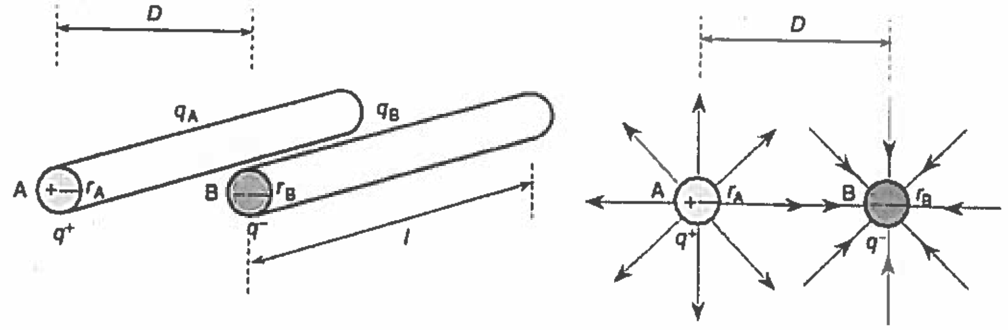
Capacitance and Capacitive Reactance:

Capacitance exists among transmission line conductors due to their potential difference. To evaluate the capacitance between conductors in a surrounding medium with permittivity ε, it is necessary to determine the voltage between the conductors, and the electric field strength of the surrounding. (4m)

Capacitance of a Single-Phase Line with Two Wires:

Consider a two-wire single-phase line with conductors A and B with the same radius r, separated by a distance D > r_A and r_B. The conductors are energized by a voltage source such that conductor A has a charge q+ and conductor B a charge q- as shown in Fig.

The charge on each conductor generates independent electric fields. Charge q+ on conductor A generates a voltage V_{AB-A} between both conductors. Similarly, charge q- on conductor B generates a voltage V_{AB-B} between conductors.



(4m)

Electric field produced from a two-wire single-phase system.

V_{AB-A} is calculated by integrating the electric field intensity, due to the charge on conductor A, on conductor B from r_A to D

$$V_{AB-A} = \int_{r_A}^{D} E_A\,dx = \frac{q}{2\pi\varepsilon_0}\ln\left[\frac{D}{r_A}\right]$$

V_{AB-B} is calculated by integrating the electric field intensity due to the charge on conductor B from D to r_B

$$V_{AB-B} = \int_{D}^{r_B} E_B\,dx = \frac{-q}{2\pi\varepsilon_0}\ln\left[\frac{r_B}{D}\right]$$

The total voltage is the sum of the generated voltages V_{AB-A} and V_{AB-B}

$$V_{AB} = V_{AB-A} + V_{AB-B} = \frac{q}{2\pi\varepsilon_0}\ln\left[\frac{D}{r_A}\right] - \frac{q}{2\pi\varepsilon_0}\ln\left[\frac{r_B}{D}\right] = \frac{q}{2\pi\varepsilon_0}\ln\left[\frac{D^2}{r_A r_B}\right]$$

If the conductors have the same radius, $r_A=r_B=r$, then the voltage between conductors V_{AB}, and the capacitance between conductors C_{AB}, for a 1-m line length are

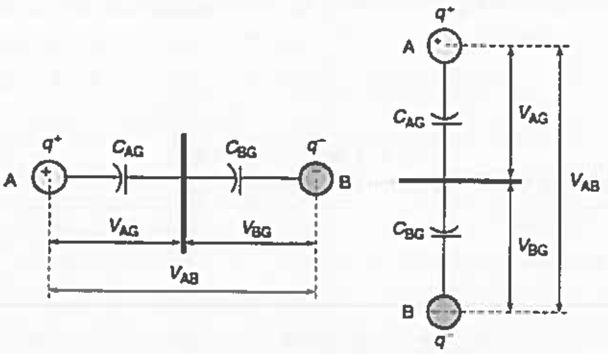
$$V_{AB} = \frac{q}{\pi\varepsilon_0}\ln\left[\frac{D}{r}\right] \text{ (V)}$$

$$C_{AB} = \frac{\pi\varepsilon_0}{\ln\left[\frac{D}{r}\right]} \text{ (F/m)}$$

The voltage between each conductor and ground (G) is one-half of the voltage between the two conductors. Therefore, the capacitance from either line to ground is twice the capacitance between lines (4m)

$$V_{AG} = V_{BG} = \frac{V_{AB}}{2} \text{ (V)}$$

$$C_{AG} = \frac{q}{V_{AG}} = \frac{2\pi\varepsilon_0}{\ln\left[\frac{D}{r}\right]} \text{ (F/m)}$$



12. Classify the types of transmission lines with model representations.

Classification of Transmission Lines - Short, Medium & Long Transmission Lines:

A. AC transmission line, and
B. DC transmission line

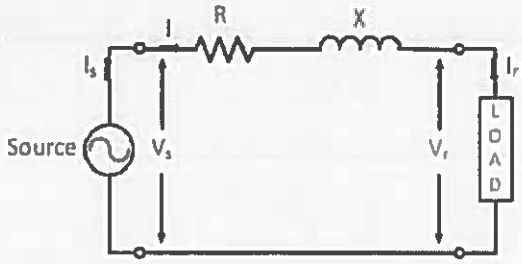
Depending upon the operating voltage and length, the overhead ac transmission lines are classified as,

1. Short transmission lines
2. Medium transmission lines
3. Long transmission lines

Short Transmission Line:

- ✓ Overhead transmission line is less than 50km
- ✓ Operating voltages of less than 20kV.

In these lines, the effect of capacitance is neglected due to smaller length and low operating voltage. Hence, the resistance and inductance effects of the line are considered while determining the performance of the short transmission line as shown below.



Equivalent circuit model of a short line

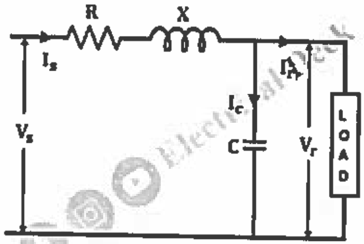
Circuit Globe

Medium Transmission Line:

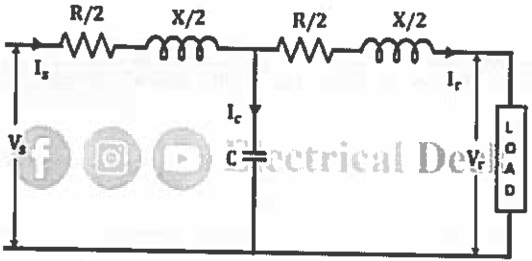
- ✓ Length of the overhead transmission line is in the range of 50-150km
- ✓ the operating voltage is greater than 20kV.

Based on the location of the capacitance at different places, the medium transmission lines have different configurations. These configurations show the different ways in which the effect of capacitance is taken into consideration. The three configurations based on the location of capacitance are,

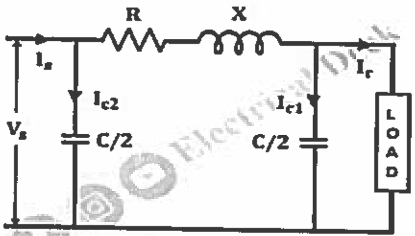
i. End condenser representation of medium transmission lines
ii. Nominal-T representation of medium transmission lines
iii. Nominal-π representation of medium transmission lines.



End Condenser Method of Medium Transmission Line



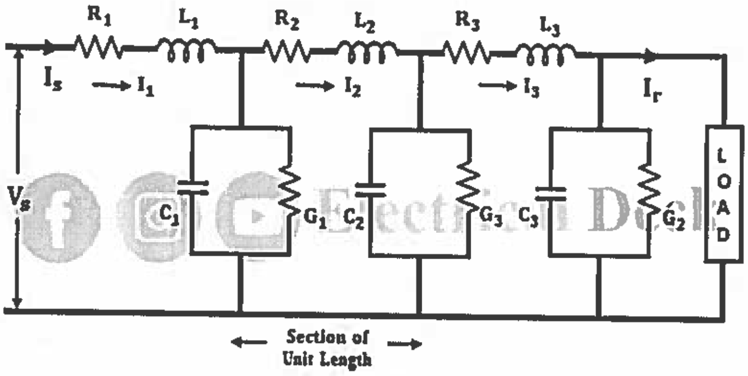
Nominal-T Method of Medium Transmission Line



Nominal-π Method of Medium Transmission Line

Long Transmission Line:

- ✓ Overhead transmission lines whose length is more than 150km
- ✓ Operating voltage of these lines is more than 100kV



Long Transmission Line

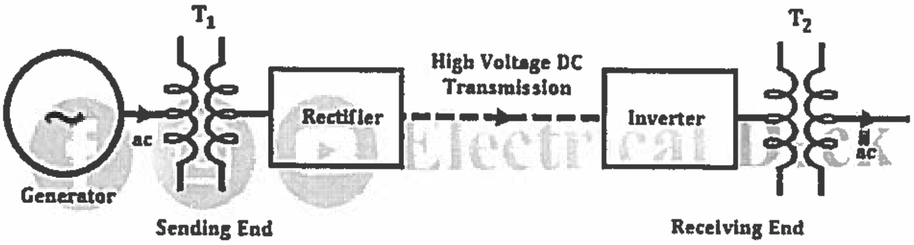
DC Transmission Line

In AC transmission lines, for transmission of power over long distances at higher voltages, the cost of transmission line and loss increases.

Also, the AC long transmission line suffers from problems like stability limits, voltage control, line compensation, interconnection of lines, ground impedance, etc due to an increase in voltage levels and distance.

The various problems associated with long-distance AC transmission have led to the development of HVDC (high voltage direct current) transmission nothing but a dc transmission line.

The use of DC power for long transmission lines has various advantages like no stability problem, absence of charging current, no skin effect, need for reactive compensation, bulk power transfer, economic power transmission, etc.



HVDC Transmission System

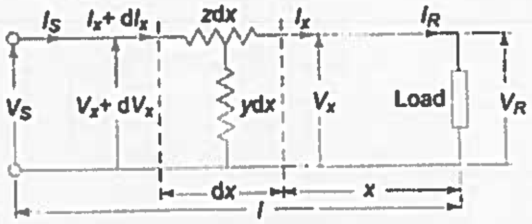
It is inefficient to use a dc transmission system for shorter and medium distances.

13. (a) Derive the expressions for the Performance of long transmission lines using rigorous method with relevant equations.

ABCD parameters in Long Transmission Line-rigorous Solution:

For rigorous solution, consider the Line Impedance and admittance uniformly distributed and not lumped.

Consider a small element of length dx situated at a distance x from receiving end. Let Z and Y denoted respectively the Series Impedance and shunt Admittance of the line per unit length.



Schematic diagram of a long line

The voltage at two ends of the element are denotes as V (towards receiving end) and V + dV (towards sending end) respectively.

Let dx be an elemental section of the line at a distance x from the receiving-end having a series impedance zdx and a shunt admittance ydx. The rise in voltage to neutral over the elemental section in the direction of increasing x is dV_x.

We can write the following differential relationships across the elemental section:

$$dVx = I_x z dx \text{ or } \frac{dV_x}{dx} = zI_x \qquad (5.14)$$

$$dIx = V_x y dx \text{ or } \frac{dI_x}{dx} = yVx \qquad (5.15)$$

It may be noticed that the kind of connection (e.g. T or π) assumed for the elemental section, does not affect these first order differential relations. Differentiating Eq. (5.14) with respect to x, we obtain

$$\frac{d^2V_x}{dx^2} = \frac{dI_x}{dx} z$$

Substituting the value of dI_x/dx from Eq. (5.15), we get

$$\frac{d^2V_x}{dx^2} = yzV_x \qquad (5.16)$$

This is a linear differential equation whose general solution can be written as follows:

$$V_x = C_1 e^{\gamma x} + C_2 e^{-\gamma x} \qquad (5.17)$$

where

$$\gamma = \sqrt{yx} \qquad (5.18)$$

and C_1 and C_2 are arbitrary constants to be evaluated.

Differentiating Eq. (5.17) with respect to x:

$$\frac{dV_x}{dx} = C_1 \gamma e^{\gamma x} - C_2 \gamma e^{-\gamma x} = zI_x$$

$$I_x = \frac{C_1}{Z_c} e^{\gamma x} - \frac{C_2}{Z_c} e^{-\gamma x} \qquad (5.19)$$

Where

$$Z_c = \left(\frac{z}{y}\right)^{1/2} \qquad (5.20)$$

The constants C_1 and C_2 may be evaluated by using the end conditions, i.e. when x = 0, $V_x = V_R$ and $I_x = I_R$. Substituting these values in Eqs. (5.17) and (5.19) gives

$$V_R = C_1 + C_2$$

$$I_R = \frac{1}{Z_c}\ (C_1 - C_2)$$

which upon solving yield

$$C_1 = \frac{1}{2}\ (V_R + Z_cI_R)$$

$$C_2 = \frac{1}{2}(V_R - Z_cI_R)$$

With C_1 and C_2 as determined above, Eqs. (5.17) and (5.19) yield the solution for V_x and I_x as

$$V_x = \left(\frac{V_R + Z_cI_R}{2}\right)e^{\gamma x} + \left(\frac{V_R - Z_cI_R}{2}\right)e^{-\gamma x}$$

$$I_x = \left(\frac{V_R/Z_c + I_R}{2}\right)e^{\gamma x} - \left(\frac{V_R/Z_c - I_R}{2}\right)e^{-\gamma x} \qquad (5.21)$$

Here Z_c is called the **characteristic impedance** of the Long Transmission Line and γ is called the **propagation constant.**

Knowing V_R, I_R and the parameters of the line, using Eq. (5.21) complex number rms values of V_x and I_x at any distance x along the line can be easily found out.

A more convenient form of expression for voltage and current is obtained by introducing hyperbolic functions. Rearranging Eq. (5.21), we get

$$V_x = V_R\left(\frac{e^{\gamma x} + e^{-\gamma x}}{2}\right) + I_RZ_c\left(\frac{e^{\gamma x} - e^{-\gamma x}}{2}\right)$$

$$I_x = V_R\frac{1}{Z_c}\left(\frac{e^{\gamma x} - e^{-\gamma x}}{2}\right) + I_R\left(\frac{e^{\gamma x} + e^{-\gamma x}}{2}\right)$$

These can be rewritten after introducing hyperbolic functions, as

$$V_x = V_R \cosh\ \gamma x + I_RZ_c \sinh\ \gamma x \qquad (5.22)$$

$$I_x = I_R \cosh\ \gamma x + V_R\ \frac{1}{Z_c}\sinh\ \gamma x$$

where x =*l*, $V_x = V_s$, $I_x = I_s$

$$\therefore \qquad \begin{bmatrix} V_S \\ I_S \end{bmatrix} = \begin{bmatrix} \cosh \gamma l & Z_c \sinh \gamma l \\ \frac{1}{Z_c}\sinh \gamma l & \cosh \gamma l \end{bmatrix} \begin{bmatrix} V_R \\ I_R \end{bmatrix} \qquad (5.23)$$

Here

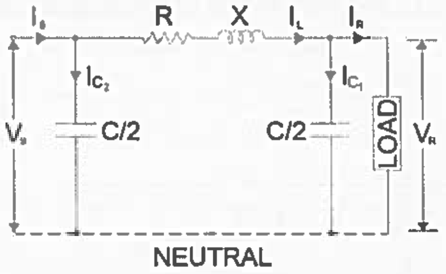
$$A = D = \cosh\ \gamma l$$

$$B = Z_c \sinh\ \gamma l \qquad (5.24)$$

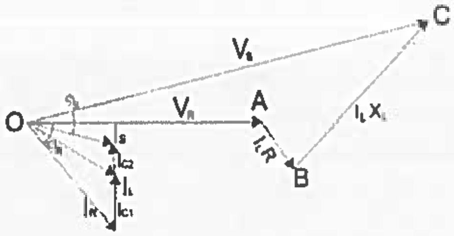
$$C = \frac{1}{Z_c}\sinh\ \gamma l$$

13. (b) Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.

In Nominal π Method, the shunt capacitance of each line i.e. phase to neutral is divided into two equal parts. One part is lumped at the sending end while the other is lumped at receiving end as shown in figure below.



Notice that, in this method there is no effect of shunt capacitance at sending end on the line voltage drop and hence on voltage regulation but this accounts for the charging current in sending end.
Let
IR = Load Current per phase
R = Resistance per phase
X = Reactance per phase
C = Capacitance per phase
Cos Ø R = Receiving end power factor (lagging)
Vs = Sending end voltage
Let us now draw the phasor. Assume receiving end voltage VR as reference and load current IR lagging this voltage by ØR.



Therefore,
VR= VR+ j0
and IR= IR<-ØR = IR (CosØR – jSinØR)
Charging Current at load end IC1 = j(ωC/2) VR = jπfCVR
Line Current **IL = IR + IC1** (phasor sum)
Sending end voltage Vs = VR + IL (R+jX)
Now,
Charging current at sending end IC2 = j(ωC/2)Vs = jπfCVs
Hence, Sending end current **Is = IL + IC2** (phasor sum)
Thus, sending end current and voltage is calculated as above and from these parameters the performance of line is evaluated.

14. Explain the different methods used to improve the string efficiency of insulators with necessary diagrams.

Methods of Improving String Efficiency

Method # 1. By Using Insulators with Larger Discs or by Providing Each Insulator Unit with a Metal Cap:

It is clear from the expression of string efficiency that the string efficiency increases with the decrease in value of K (i.e. the ratio of shunt capacitance to mutual capacitance). One method is to design the units such that the mutual capacitance (capacitance of each unit) is much greater than the shunt capacitance (capacitance to earth). This can be achieved by using insulators with larger discs or providing each insulator unit with a metal cap. The ratio K can be made 1/6 to 1/10 by this method.

Method # 2. By Using Longer Cross-Arms:

The ratio of shunt capacitance to mutual capacitance, K can alternatively be reduced by using longer cross-arms so that the horizontal distance from line support (pole or tower) is increased thereby decreasing the shunt capacitance. But the limitations of cost and mechanical strength of line supports do not allow the cross-arms to be too long and it has been found that in practice it is not possible to obtain the value of K less than 0.1.

Method # 3. By Capacitance Grading:

It is seen that non-uniform distribution of voltage across an insulator string is due to leakage current from the insulator pin to the supporting structure, which cannot be eliminated. However, it is possible that discs of different capacities are used such that the product of their capacitive reactance and the current flowing through the respective unit is same. This can be achieved by grading the mutual capacitance of the insulator units i.e., by having lower units of more capacitance—maximum at the line unit and minimum at the top unit, nearest to the cross-arm. By this method complete equality of voltage across the units of an insulator string can be obtained but this method needs a large number of different-sized insulator units and maintaining spares of all varieties of insulator discs. So this method is not used in practice below 200 kV.

Consider a 4-unit string. Let C be the capacitance of the top unit and let the capacitances of others units are C_2, C_3 and C_4, as shown in Fig.

Assume $C_1 = k\,C$

Applying Kirchhoff's first law to node A we get, $I_2 = I_1 + i_1$

$\Rightarrow \omega C_2 v = \omega C v + \omega C_1 v$ or $C_2 = C + K C = C(1 + K)$ ... (9.11)

Applying Kirchhoff's first law to node B we get, $I_3 = I_2 + i_2$

or $\omega C_3 v = \omega C_2 v + \omega C_1 \times 2v$

(or) $C_3 = C_2 + 2KC = C(1 + K) + 2KC = C(1 + 3K)$ ... (9.12)

Applying Kirchhoff's first law to node C we get,

$I_4 = I_3 + i_3$

or $\omega C_4 v = \omega C_3 v + \omega C_1 \times 3v$

or $C_4 = C_3 + 3KC = C(1 + 3K) + 3KC = C(1 + 6K)$ ...(9.13)

Thus, it will be possible to equalize the potential across all the units, if their capacitances are in the ratio of 1: (1 + K): (1 + 3 K): (1 + 6 K) and so on.

But in practice it is impossible to obtain such units which will have their capacitances in above ratio, although nearby results can be obtained by employing standard insulators for most of the units and employing larger units adjacent to the line.

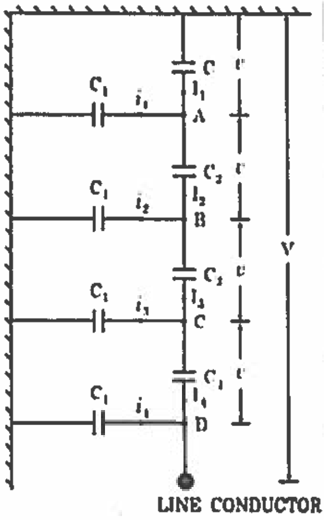
Method # 4. By Static Shielding:

In case of capacitance grading, insulator units of different capacitances are used so that the flow of different currents through the respective units produce equal voltage drop. In static shielding, pin to supporting structure charging currents are exactly cancelled so that the same current flows through the identical insulator units and produce equal voltage drops across each insulator unit.

This arrangement is given is Fig. 9.22. In this method a guard or grading ring, which usually takes the form of a large metal ring surrounding the bottom unit and electrically connected to the metal work at the bottom of this unit, and therefore to the line conductor.

The guard ring screens the lower units, reduces their earth capacitance C_1 and introduces a number of capacitances between the line conductor and the various insulator unit caps. These capacitances are greater for lower units and thus the voltages across them are reduced. With this method also it is impossible to obtain in practice an equal distribution of voltage but considerable improvements are possible.

Let the capacitances between the links and the shield be C_x, C_y and C_z respectively as shown in Fig. 9.22, and let v be the potential across each unit.



Since the capacitance of each unit is same, therefore, their charging currents I_1, I_2, I_3, and I_4 would be same, let it be I. If $C_1 = K C$

Applying Kirchhoff s first law to node A we get,

$I + i_x = I + i_1$ or $i_x = i_1$... (9.14)

Similarly, $i_y = i_2$... (9.15)

and $i_z = i_3$ ... (9.16)

Also, $i_1 = \omega C_1 v = \omega K C v$ ... (9.17)

$i_2 = 2 \omega C_1 v = 2 \omega K C v$ ... (9.18)

$i_3 = 3 \omega C_1 v = 3 \omega K C v$ ... (9.19)

The potential causing current i_x is 3 v (voltage across three units leaving the top one).

So, $i_x = \omega C_x \times 3 v = 3 \omega C_x v$ ... (9.20)

Comparing Eqs. (9.14), (9.17) and (9.20), we have,

$3 \omega C_x v = \omega K C v$ or $C_x = K C/3$ ... (9.21)

The potential causing current y is 2 v and therefore,

$I_y = 2 \omega C_y v$ ... (9.22)

Comparing Eqs. (9.15), (9.18) and (9.22) we have,

$2 \omega C_y v = 2 \omega K C v$ or $C_y = K C$ ... (9.23)

The potential causing current i_z is v and therefore,

$i_z = \omega C_z v$ ... (9.24)

Comparing Eqs. (9.16), (9.19) and (9.24), we have,

$\omega C_z v = 3 \omega K C v$

or $C_z = 3 K C$

In general, if there are n units

$i_1 = \omega K C v$ and $i_x = (n-1) \omega C_x v$

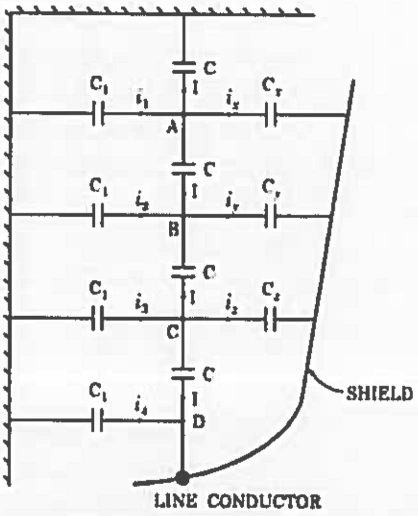
or $C_x = KC/ (n-1)$

Similarly, $C_y = 2KC/ (n-2)$

and $C_z = 3KC/ (n-3)$

or The capacitance of the pth metal link to the line is given as:

$C_p = pKC/ (n-p)$

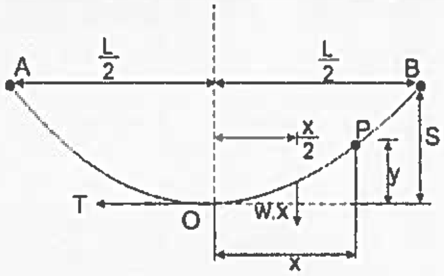


15. (a) Derive an expression for sag of a line supported between two supports of the same tower height.

Calculation of Sag in a Transmission Line:

Sag calculation for supports is at equal levels

Suppose, AOB is the conductor. A and B are points of supports. Point O is the lowest point and the midpoint.



Let, L = length of the span, i.e. AB

w = weight per unit length of the conductor

T = tension in the conductor.

We have chosen any point on the conductor, say point P.

The distance of point P from the Lowest point O is x.

y is the height from point O to point P.

Equating two moments of two forces about point O as per the figure above we get,

$$Ty = wx \times \frac{x}{2}$$

$$Now, y = \frac{wx^2}{2T},$$

The maximum dip (sag) is represented by the value of y at either of the supports A and B. At support A, x = l/2 and y = S.

$$Then\ S = \frac{wL^2}{8T}$$

15. (b) A 132 kV transmission line has the following data:
Wt. of conductor = 680 kg/km;
Length of span = 260 m
Ultimate strength = 3100 kg;
Safety factor = 2.
Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 m.

Sol: Weight of conductor/mt run, $w = \frac{680}{1000} = 0.68$ kg.

Working Tension, $T = \frac{\text{Ultimate Strength}}{\text{Safety factor}}$

$= \frac{3100}{2} = 1550$ kg.

Span length, $l = 260$ m.

$\therefore$ Sag $= \frac{wl^2}{8T} = \frac{0.68 \times 260^2}{8 \times 1550}$

$= 3.7$ m.

Height = GC + Sag = 10 + 3.7
= 13.7 m.

UshaRani
(Faculty:)

7/11/23.
(HOD)

Semester End Regular/Supplementary Examination, Dec./Jan., 2022-2023

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2022 - 2023
Course Code	20EC305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Digital System Design**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Draw and write the truth table for EXOR gate.	20EC305.1	L1
2	Give an example for Min term and Max term involving 3 Variables.	20EC305.2	L1
3	What is Carry Dependency?	20EC305.3	L1
4	What is Race round condition?	20EC305.4	L1
5	Expand VHDL.	20EC305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Convert the following (i) $ABC_{16}=(\ )_{10}$ (ii) $154_8=(\ )_{10}$ (iii) $695_{10}=(\ )_{16}$.	6M	20EC305.1	L2
6 (b)	Perform the given subtraction using 1's and 2's complement methods: $(11100011)_2-(01110000)_2$.	6M	20EC305.1	L2
	OR			
7	Explain basic gates with their truth table.	12M	20EC305.1	L2
8	Implement AND gate, OR gate, EXOR gate and Not gate using NAND gate only.	12M	20EC305.2	L2
	OR			
9 (a)	Simplify the expression Y=(A+B) (A'+C)(B'+C').	6M	20EC305.2	L2
9 (b)	Minimize the following function using Karnaugh map technique $f(A,B,C,D)= \Sigma_m(5,6,7,12,13)+ \Sigma_d(4,9,14,15)$	6M	20EC305.2	L2
10 (a)	Design the full adder using two half adders.	6M	20EC305.3	L3
10 (b)	Design a 4-bit BCD adder circuit.	6M	20EC305.3	L3
	OR			
11 (a)	Differentiate PAL, PROM, and PLA.	6M	20EC305.3	L2
11 (b)	Show and implement the following function using a PROM $F(w,x,y,z)= \Sigma_m(1,9,11,12,13,15)$ $G(w,x,y,z)= \Sigma_m(0,1,2,3,4,5,7,8,10,11,12,13,14,15)$	6M	20EC305.3	L3
12 (a)	Explain the working of JK Flip Flop.	6M	20EC305.4	L2
12 (b)	Explain the Conversion of SR Flip Flop to T Flip Flop.	6M	20EC305.4	L2
	OR			
13	Design a 3 bit up counter using D-flip flop.	12M	20EC305.4	L3
14	Explain the program structure of VHDL and Explain the significance of entity and architecture.	12M	20EC305.5	L2
	OR			
15	List and discuss various data types in VHDL with examples.	12M	20EC305.5	L2

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(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

Degree- B.Tech - ECE
course Code- 20EC305
PART-A

ANSWER KEY AND SCHEME OF EVALUATION
DIGITAL SYSTEM DESIGN

AC : 2022-2023
SEM : III

1. Draw and write the truth table for EXOR gate (2M)
   XOR gate Symbol – 1m
   truth table – 1m

2. Give an example for min and Max term involving 3 Variables (2m)
   Minterm example – 1m
   Maxterm example – 1m

3. what is carry dependency – 2m
   Carry dependency – 2m

4. what is race around Condition – 2m
   race around Condition – 2m

5. Expand VHDL
   VHDL abrevation – 2m

PART-B

6a) Convert the following 6m
i) $ABC_{16} = ()_{10}$ – 2m ii) $154_8 = ()_{10}$ – 2m iii) $695_{10} = ()_{16}$ – 2m

6(b) Perform the given subtraction using 1's & 2's Compliment methods – 6m
1's Compliment method – 3m
2's Compliment method – 3m

7) Explain Basic gates with their truth table
AND gate – 4m
OR gate – 4m
NOT gate – 4m
(Symbol – 1m, expression – 1m, explaination – 1m, truth table – 1m)

8. Implement AND gate, OR gate, EXOR gate, and NOT gate using NAND gates only.
AND gate realisation – 3m
OR gate realisation - 3m
EXOR gate realisation – 4m
NOT gate realization – 2m

9(a) Simplify the expression $Y=(A+B)(\bar{A}+C)(\bar{B}+\bar{C})$
$Y=(A+B)(\bar{A}+C)(\bar{B}+\bar{C})$ simplification – 6m

9(b) minimize the following using k-map technique
$f(A,B,C,D)=\Sigma_m(5,6,7,12,13)+\Sigma_d(4,9,14,15)$ → (6m)

10 a) Design the full adder using two half adder
full adder truth table – (2m)
design a full adder – (4m)

10 b) Design a 4 bit BCD adder circuit.
Truth table for BCD adder circuit –(2m)
expressions. (2m)
Circuit (2m)

11(a) Differentiate PAL, PROM, and PLA (6m)
any four differences (6m)

11(b) Show and implement the following functions using PROM
implementation (6m)

12(a) explain the working of JK flip flop (6m)
Circuit diagram (2m) Explaination (1m) Truth table (1m)
characteristic table (1m) excitation table (1m)

12(b) explain Conversion of SR FF to T FF. – 6m
characterstic table – T – 1m simplification – 1m
excitation table – SR – 1m Circuit – 2m
Table for SR & T – 1m

13(a) Design a 3bit up Counter using D-ff – 12m
state table – 2m, Circuit diagram – 4m, timing diagram – 4m
explaination – 2m.

14. explain the program structure of VHDL & explain the significance of entity & architecture – 12m.
Program structure – 2m, entity syntax - 3m, Architecture syntax - 3m
significance – 4m

15. list and discuss various data types in VHDL with ~~explaina~~ exam
(12
data types list (8m) each data type explaination –(9m)

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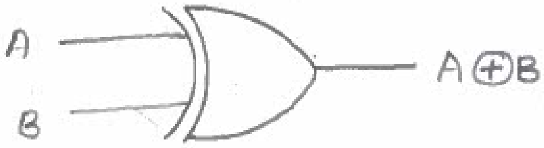
ANSWER KEY AND SCHEME OF EVALUATION

PART- A:

1. Draw and write the truth table for EXOR gate.

A.



A	B	A⊕B
0	0	0
0	1	1
1	0	1
1	1	0

2. Give an example for Min term and Max term involving 3 Variable.

A. Min term: $f(A,B,C) = (\Sigma m(1,3,5,7)$

$Y = \bar{A}\bar{B}C + \bar{A}BC + A\bar{B}C + ABC$

max term: $f(A,B,C) = \pi_m(1,3,5,7)$

$Y = (A+B+\bar{C})(A+\bar{B}+\bar{C})(\bar{A}+B+\bar{C})(\bar{A}+\bar{B}+\bar{C})$

3. what is carry dependency?

A. In Adder circuit next stage circuit will depends on previous stage carry. To generate the next stage output it has to wait for previous stage carry. This dependency is called as carry dependency.

4) What is Race around Condition?

A) race around Condition in JK flipflop, if J=1 and K=1 and if Clk=1 for a long period of time, then Q output will toggle as long as CLK is high, which makes the output of the flip-flop unstable or uncertain. This problem is called race around condition.

5) Expand VHDL?

VHDL — VHSIC HDL

Very high speed integrated circuit hardware description language.

6a Convert the following (i) $ABC_{16} = ()_{10}$ ii) $154_8 = ()_{10}$ iii) $695_{10} = ()_{16}$

Sol (i) $ABC_{16} = ()_{10}$

By using power method

$$\begin{array}{ccc} A & B & C \\ 16^2 & 16^1 & 16^0 \end{array}$$

$16^2(A) + 16^1(B) + 16^0(C)$

$256(10) + 16(11) + 1(12)$

$2560 + 176 + 12$

$(2748)_{10}$

ii) $(154)_8 = ()_{10}$

By using power method

$$\begin{array}{ccc} 1 & 5 & 4 \\ 8^2 & 8^1 & 8^0 \end{array}$$

$8^2(1) + 8^1(5) + 8^0(4)$

$64 + 40 + 4$

$(108)_{10}$

(iii) $695_{10} = ()_{16}$

By Successive division method

16 | 695
16 | 43 - 7
  2 - 11

$11 \rightarrow B$

$(2B7)_{16}$

6b Perform the given Subtraction Using 1s and 2's Compliment methods : $(11100011)_2 - (01110000)_2$.

A: $(11100011)_2 - (01110000)_2$

01110000

10001111 ← 1's Compliment

11100011

\+ 10001111

$\boxed{1}$01110010

There is carry so positive number

add carry to LSB

01110010

\+ 1

01110011

$\therefore (11100011)_1 - (01110000)_1 = (01110011)_1$

$(11100011)_2 - (01110000)_2$

01110000

10001111

\+ 1

10010000 ← 2's Compliment

11100011

10010000

$\boxed{1}$01110011

Positive number

add carry to LSB

01110011

\+ 1

01110100

$(11100011)_2 - (01110000)_2 = (01110100)_2$

7. Explain basic gates with their truth tables

A: 1. **AND Gate :-**

An AND gate is an electrical circuit that combines two signals so that the output is on if both signals are present. The output of the AND gate is connected to the base driver which is coupled to the bases of transistors, and alternately switches the transistors at opposite corners of the inverter.

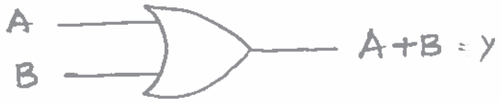
logic expression for AND gate is

$Y = A \cdot B$ Y = output variable
A, B are input variable

Truth table :-

A	B	Y = AB
0	0	0
0	1	0
1	0	0
1	1	1

2. **OR gate :-**



An OR gate is a digital logic gate that gives an output of 1 when any of its inputs are 1. Otherwise 0. An OR gate performs like two switches in parallel supplying a light, so that when either of the switches is closed the light is on. OR gates can have more than two inputs.

logic expression for OR gate is

$Y = A + B$

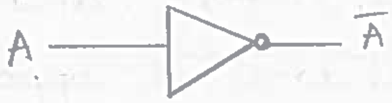
Y = output Variable

A, B are input Variables

Truth table :-

A	B	Y=A+B
0	0	0
0	1	1
1	0	1
1	1	1

NoT gate :-



The NOT gate is an electronic circuit that produces an inverted version of the input at its output. It is also known as an inverter. If the input variable is A, the inverted output is known as NOT A. This is also known as $\bar{A}$ or A'..

logic expression for NOT gate is

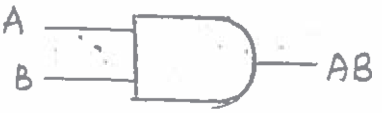
~~$X = \bar{A}$~~ $Y = \bar{A}$

Truth table :-

A	$Y=\bar{A}$
0	1
1	0

8. Implement AND gate, OR gate, EXOR gate and NOT gate Using NAND gate only. 5

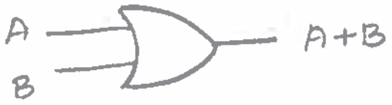
A. i. AND gate Using NAND gate

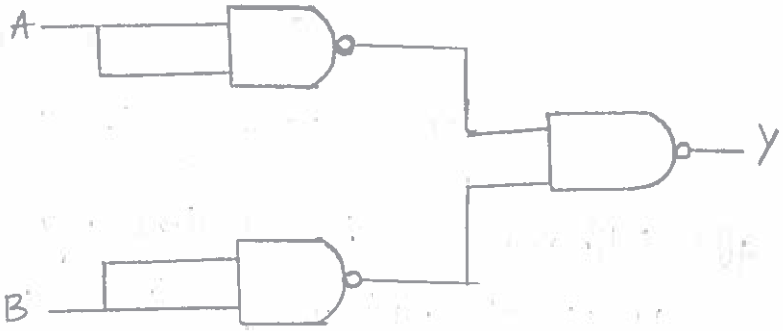


$$Y = \overline{\overline{AB}\cdot\overline{AB}}$$

$$Y = AB$$

ii. OR gate using NAND gate



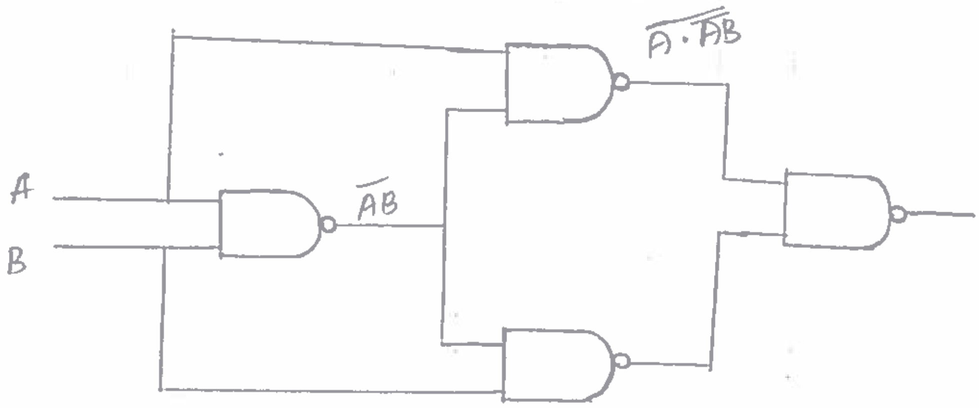


$$Y = \overline{\bar{A}\cdot\bar{B}}$$

$$Y = \bar{\bar{A}} + \bar{\bar{B}}$$

$$Y = A+B$$

iii. EXOR gate Using NAND gate



$$Y = \overline{A\cdot\overline{AB}}\cdot\overline{B\cdot\overline{AB}}$$

$$= \overline{\overline{A\cdot\overline{AB}}} + \overline{\overline{B\cdot\overline{AB}}}$$

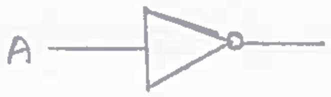
$$= A\cdot\overline{AB} + B\cdot\overline{AB}$$

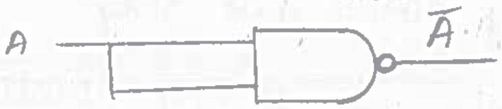
$$= A\,(\bar{A}+\bar{B}) + B\,(\bar{A}+\bar{B})$$

$$= A\bar{A} + A\bar{B} + \bar{A}B + B\bar{B}$$

$$Y = A\bar{B} + \bar{A}B = A \oplus B$$

iv. NOT gate using NAND gate.





$$\overline{AA} = \bar{A}$$

9(a) Simplify the expression $Y = (A+B)(A'+C)(B'+C')$.

A. $Y = (A+B)(A'+C)(B'+C')$

$= (AA' + A'B + AC + BC)(B'+C')$

$= AA'B' + A'BB' + AB'C + BB'C + AA'C' + A'BC' + ACC' + BCC'$

$= 0+0 + AB'C + 0 + 0 + A'BC' + 0 + 0 \quad (\because A\bar{A} = 0)$

$\therefore Y = AB'C + A'BC'$

9(b) Minimize the following function using Karnaugh map technique $f(A,B,C,D) = \Sigma_m(5,6,7,12,13) + \Sigma_d(4,9,14,15)$

A. Given $f(A,B,C,D) = \Sigma_m(5,6,7,12,13) + \Sigma_d(4,9,14,15)$

CD \ AB	$\bar{A}\bar{B}$	$\bar{A}B$	AB	$A\bar{B}$
$\bar{C}\bar{D}$	0	1	3	2
$\bar{C}D$	4 X	5 1	7 1	6 1
CD	12 1	13 1	15 X	14 X
$C\bar{D}$	8	9 X	11	10

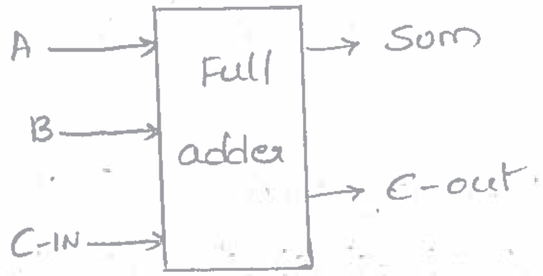
$\cancel{\bar{C}}D \ \cancel{C}D \ \cancel{\bar{A}}\cancel{\bar{B}} \ \cancel{\bar{A}}\cancel{B} \ \cancel{A}\cancel{B} \ \cancel{A}\cancel{\bar{B}}$

$Y = D$

$\therefore f(A,B,C,D) = D$ //

10a Design the full adder using two half adders. 6

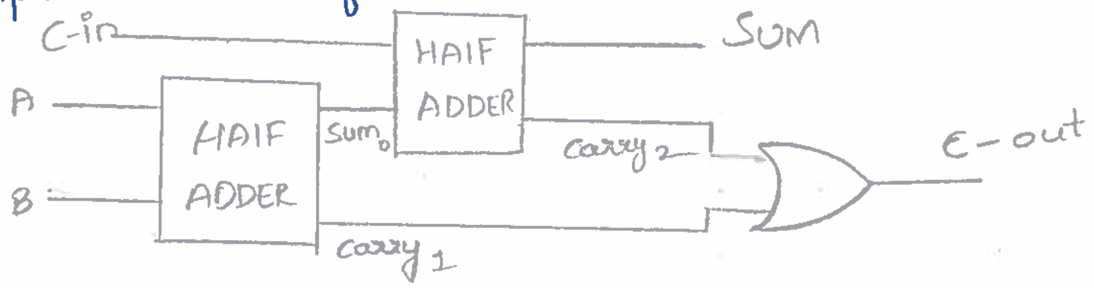
A. Full adder is the adder that adds three inputs and produces two outputs. The first two inputs are A and B and third input is input carry as C-IN. The output carry is designated as c-out and the normal output is designated as S which is SUM.



INPUTS			OUTPUTS	
A	B	C-IN	Sum	c-out
0	0	0	0	0
0	0	1	1	0
0	1	0	1	0
0	1	1	0	1
1	0	0	1	0
1	0	1	0	1
1	1	0	0	1
1	1	1	1	1

full Adder using two half adders:

2 Half adders and an OR gate is required to implement a full adder.



With this logic circuit, two bits can be added together taking carry from the next lower order of magnitude and sending a carry to the next higher order of magnitude.

10b Design a 4-bit BCD adder Circuit

A: BCD adder refers to a 4-bit binary adder that can add two 4-bit words of BCD format. The outpu of the addition is a BCD format 4-bit output word, which defines the decimal sum of the addend & augend and a carry that is created in case this sum exceeds a decimal value of 9.

Decimal Value	Binary sum					BCD Sum				
	C_3^*	S_3^*	S_2^*	S_1^*	S_0^*	C	S_3	S_2	S_1	S_0
0	0	0	0	0	0	0	0	0	0	0
1	0	0	0	0	1	0	0	0	0	1
2	0	0	0	1	0	0	0	0	1	0
3	0	0	0	1	1	0	0	0	1	1
4	0	0	1	0	0	0	0	1	0	0
5	0	0	1	0	1	0	0	1	0	1
6	0	0	1	1	0	0	0	1	1	0
7	0	0	1	1	1	0	0	1	1	1
8	0	1	0	0	0	0	1	0	0	0
9	0	1	0	0	1	0	1	0	0	1
10	0	1	0	1	0	1	0	0	0	0
11	0	1	0	1	1	1	0	0	0	1
12	0	1	1	0	0	1	0	0	1	0
13	0	1	1	0	1	1	0	0	1	1
14	0	1	1	1	0	1	0	1	0	0
15	0	1	1	1	1	1	0	1	0	1
16	1	0	0	0	0	1	0	1	1	0
17	1	0	0	0	1	1	0	1	1	1
18	1	0	0	1	0	1	1	0	0	0
19	1	0	0	1	1	1	1	0	0	1

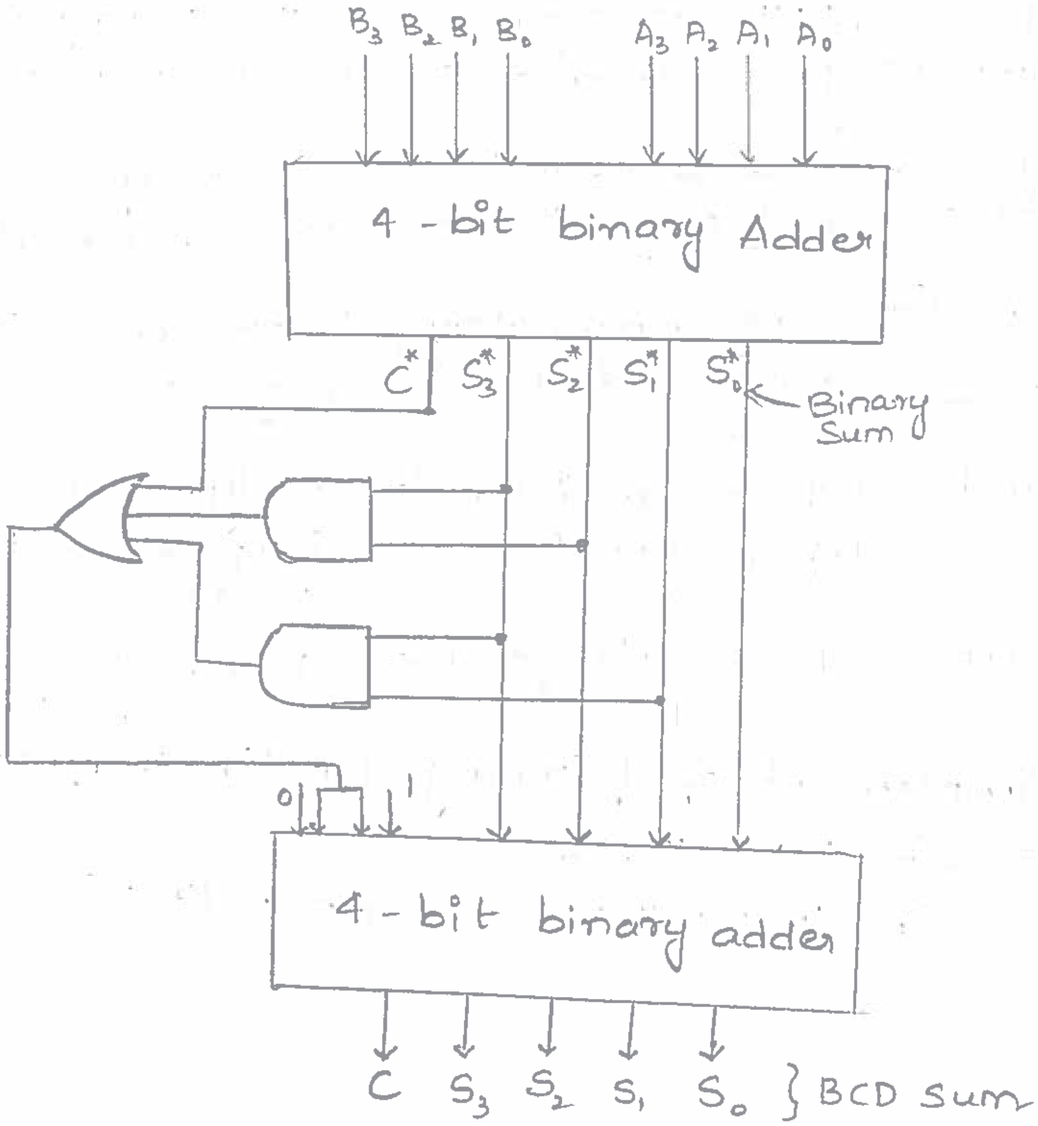
1st. Condition → $C^* = 1$

2nd. Condition → $S_3^*(S_2^* + S_1^*)$

Correction :-

$$= C^* + S_3^*(S_2^* + S_1^*)$$

$$= C^* + S_3^* S_2^* + S_3^* S_1^*$$



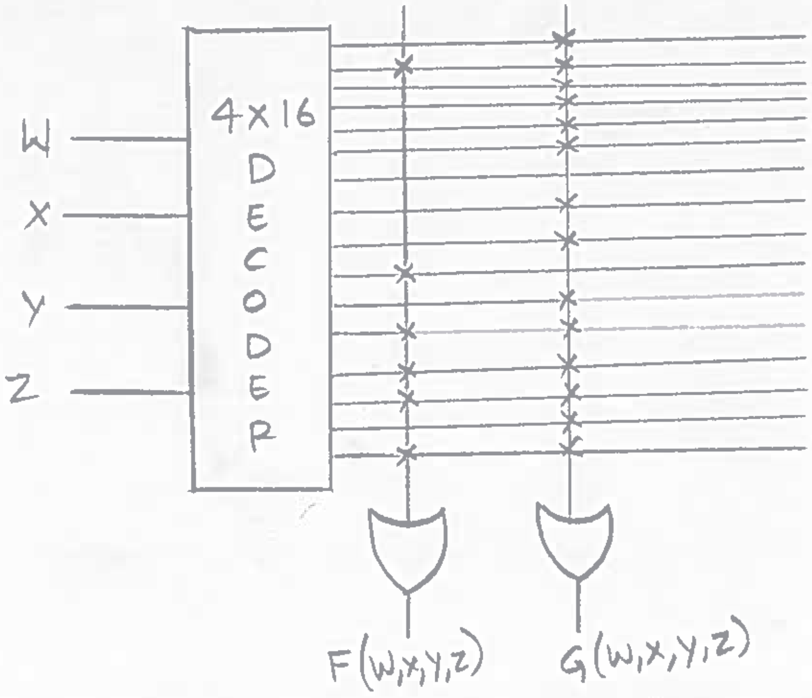
11 a. Differentiate PAL, PROM, and PLA.

PROM	PAL	PLA
1. The Decoder (AND array) implements all minterms	1. The AND array implements a limited no. of product terms.	1. The AND array implements a limited no product terms.
2. AND array is not programmable	AND array is programmable	AND array is programmable
3. OR Array is programmable	3. OR array is not programable	OR array is programmable.
4. Inter Connections are more.	Inter Connections are medium	Inter Connections are less.
5. cheap and Simple	cheap and simple than PLA.	Costly and Complex Compared to PROM & PAL.
6. least flexible	Moderate flexible	Extremely flexible

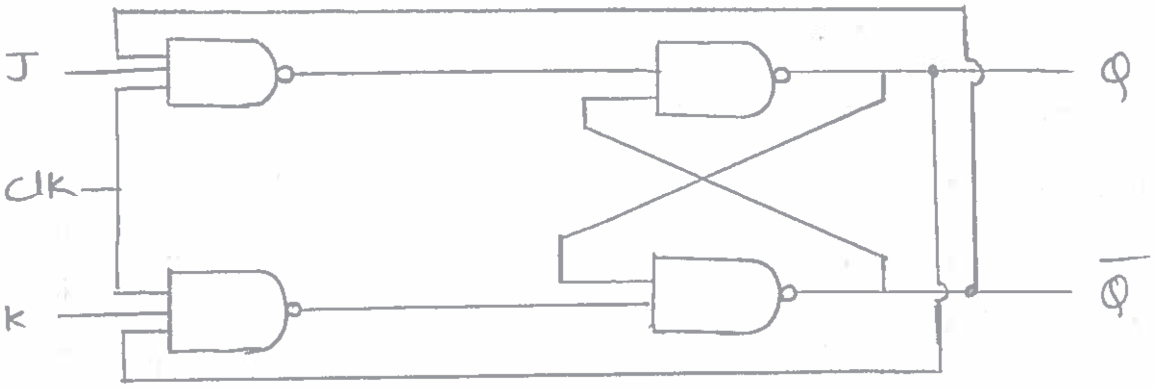
11 b: Show and implement of following function Using PROM

$F(w,x,y,z) = \sum m\,(1,9,11,12,13,15)$

$G(w,x,y,z) = \sum m\,(0,1,2,3,4,5,7,8,10,11,12,13,14,15)$



12a Explain the working of JK flip flop. 8



A J-K flip flop is nothing more than S-R flip flop with an added layer of feedback. This feedback Selectively enables one of the two set/reset inputs So that they cannot both carry on active signal to the multivibrator Circuits, thus eliminating the invalid Condition.

1. When J=0, K=0

	o/p
$Q = 0$	0
$\bar{Q} = 1$	1

when J=0, K=0 it is memory.

2. when J=0, K=1

	o/p
$Q = 0$	0
$\bar{Q} = 1$	1
$Q = 1$	0
$\bar{Q} = 0$	1

3. when J=1, K=0

	o/p
$Q = 0$	1
$\bar{Q} = 1$	0
$Q = 0$	1
$\bar{Q} = 1$	0

4. when J=1, K=1

	o/p
$Q = 0$	1
$\bar{Q} = 1$	0

	o/p
$\bar{Q} = 0$	1
$Q = 1$	0

when J=1, K=1 then the output is inverted or toggled State.

Truth table:

Clk	J	K	Q_{n+1}
0	X	X	Q_n
1	0	0	Q_n
1	0	1	0
1	1	0	1
1	1	1	$\overline{Q_n}$

excitation table.

Q_{n+1}	Q_n	J	K
0	0	0	X
0	1	X	1
1	0	1	X
1	1	X	0

Characteristic table:

Q_n	J	K	Q_{n+1}
0	0	0	0
0	0	1	0
0	1	0	1
0	1	1	1
1	0	0	1
1	0	1	0
1	1	0	1
1	1	1	0

12b) Explain the Conversion of SR flip-flop to T-flip flop.

A. Available flip flop - SR

required flip flop - T

excitation table for available flip-flop - SR

Q_n	Q_{n+1}	S	R
0	0	0	X
0	1	1	0
1	0	0	1
1	1	X	0

Characteristic table for required flip-flop - T

Q_n	T	Q_{n+1}
0	0	0
0	1	1
1	0	1
1	1	0

Combine both the tables of SR and T

Q_n	T	Q_{n+1}	S	R
0	0	0	0	x
0	1	1	1	0
1	0	1	0	1
1	1	0	x	0

Draw K-maps for S and R

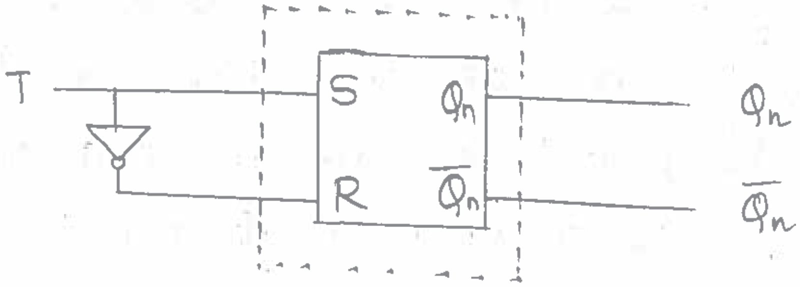
S

$Q_n \backslash T$	$\overline{T}$	T
$\overline{Q_n}$	0	1
Q_n	0	x

R

$Q_n \backslash T$	$\overline{T}$	T
$\overline{Q_n}$	X	0
Q_n	1	0

$S = T$ $R = \overline{T}$



13. Design a 3 bit up Counter using D-flip flop.

A: Consider a 3-bit counter with each bit count represented by Q_0, Q_1, Q_2 as the outputs of flipflops FF_0, FF_1, FF_2, respectively. Then the state table would be

D-flip flop updates the state according to the input applied to it i.e $Q = D$. So designing up counter using D- flip flop is different than a T- flip flop.

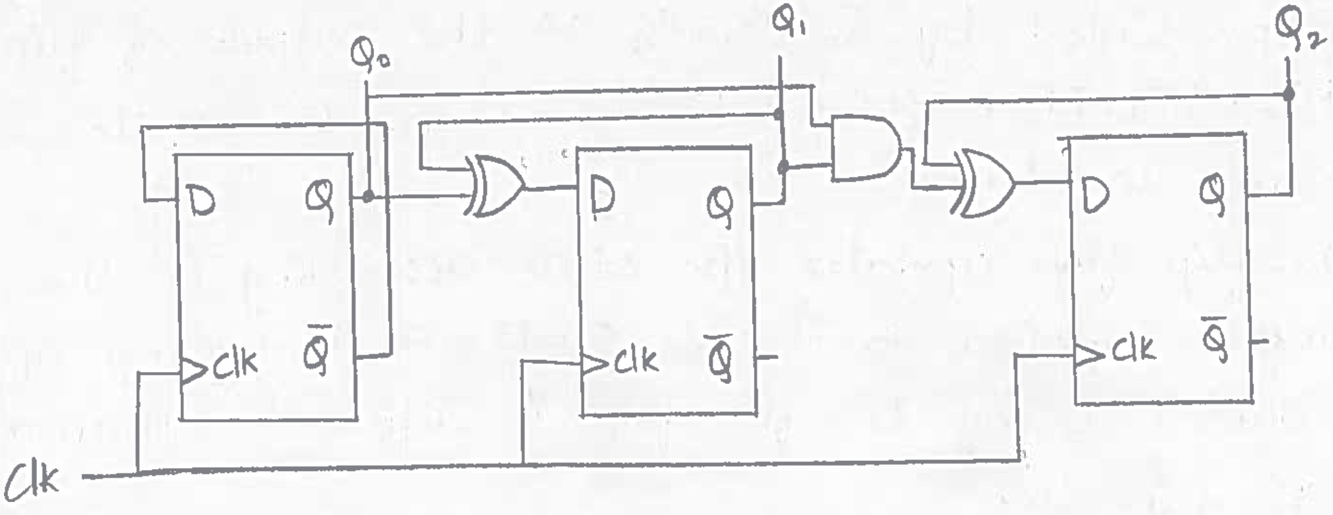
State	Q_2	Q_1	Q_0
0	0	0	0
1	0	0	1
2	0	1	0
3	0	1	1
4	1	0	0
5	1	0	1
6	1	1	0
7	1	1	1

According to the State table of up-Counter Q_0 is Continuously changing so the input to FF_0 will be $D_0 = \bar{Q}_0$. Because it will toggle the state whenev a clock pulse hits the FF_0.

$Q_1 = 1$, when its previous state Q_1 & Q_0 are not equal & $Q_1 = 0$, when its previous state Q_1 & Q_0 are equal. That is the same as XOR operation. So $D_1 = Q_1$ XOR Q

$Q_2 = 1$, when in its previous state the AND of Q_1 & Q_0 is not equal to Q_2. $Q_2 = 0$, when in its previous state th AND of Q_1 & Q_0 is equal to Q_2. So $D_2 = Q_2$ XNOR (Q_1 & Q

Schematic of synchronous up-counter using D-ff is given below.



14 Explain the program structure of VHDL and explain the significance of entity and architecture. 10

A. Program Structure of VHDL

```
library declaration ;
Entity   Entityname is
Port ( port_name : signal_mode  signal_type ;
          :
         Port_name : signal_mode  signal_type);
    END  Entity name ;
Architecture   architecture_name  Of entity_name is
  ( declarations)
 Begin
  ~~(Code)~~
~~END architecture_name ;~~
     Signal declaration
     Component declaration
      Body
  End architecture_name ;
```

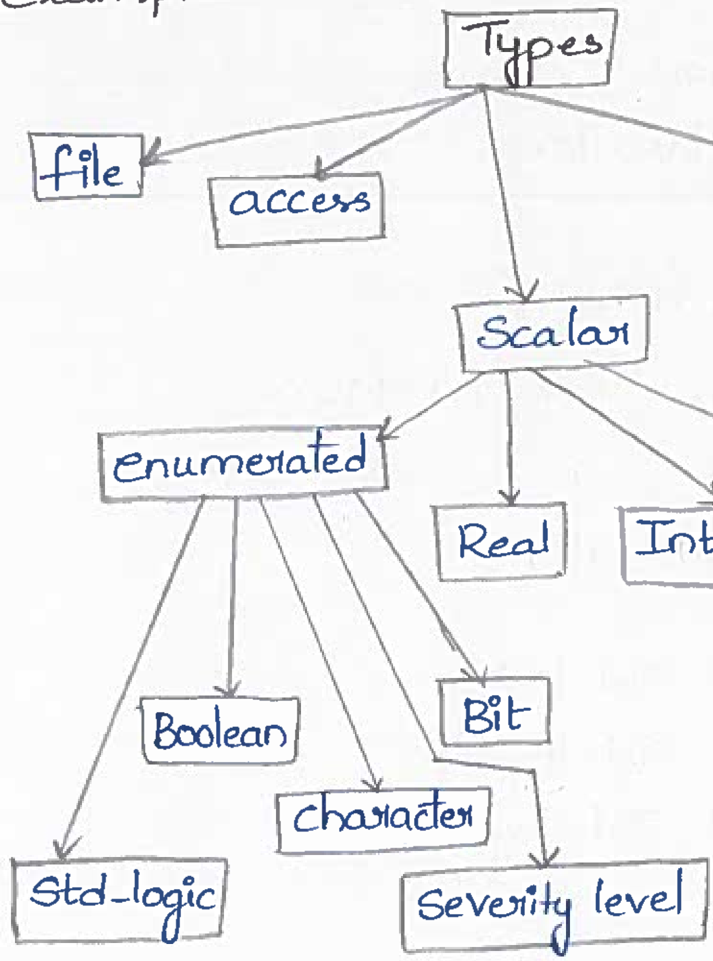
Example :- VHDL program for 4x1 MUX.

```
library ieee ;
Use ieee. std_logic_1164.all ;
entity  4x1 MUX is
 Port  ( S1 , S0  :  in  std_logic ;
         A,B,C,D  :  in  std_logic ;
             Y    :  out std_logic) ;
 end  4x1 MUX ;
```

```
architecture 4X1MUX-1 of 4X1MUX is
 Begin
  With select S1 S0 Y <= A when "00",
                         B when "01",
                         C when "10",
                         D when "11",
                         U when others;
  end 4X1 MUX-1 ;
```

A VHDL models Consist of an entity declaration and a architecture Body. The entity defines the interface, the architecture defines the function. The entity declaration names the entity and defines the interface to its environment.

15 List and discuss Various data types in VHDL, with examples.



Scalar types: Scalar types describe objects that can hold, at most, one value at a time. The type itself can contain multiple values, but an object that is declared to be a scalar type can hold, at most, one of the scalar values at any point in time.

1. Integer types
2. Real types
3. Enumerated types
4. Physical types
5. floating point

Enumerated Data types

An enumerated data type is a very powerful tool for abstract modelling. A designer can use an enumerated type to represent exactly the values required for a specific operation.

1. Boolean
2. character
3. Bit
4. Std-logic
5. Severity level

1. **Boolean:**

This data type is used when we need to convey some true or false conditions.

Eg: Architecture .....

```
Begin
Process
(....)
if a<b then
 temp <= True;
 else
temp <= False;
end if
end Process;
```

Character:

This data type is used when we need to use all alpha numeric and special characters.

Bit:

This data type is used when we need to represent binary values ('0' and '1').

Severity level

This data type is used in complex projects when we need to show warnings, errors in runtime, failure in runtime.

Std_ulogic :-

This data types are declared in std-logic-1164. all package of Ieee library.

Integer Data type:

Integers are exactly like mathematical integers. All of the normal predefined mathematical function like add, subtract, multiply, and divide apply to integer types

Eg:- 1) Type - integer declaration

type < word lengt > is range 0 to 31;

2) object - integer declaration

Constant < loop number >:< integer >< =345;

Real Data type:

Real types are used to declare objects that emulate mathematical real numbers. They can be used to represent numbers out of the range of integer values as well as fractional values.

Physical Data types:

Physical types are used to represent physical quantities such as distance, Current, time and so on. A physical type provides for a base unit and successive units are then defined in terms of this unit.

Eg: Type Current is range 0 to 10000000

units
na; -- nano amps
ua = 1000na; -- micro amps
ma = 1000 μa; -- milli amps
a = 1000 ma; -- amps
end units;

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7/1/23

Semester End Regular/Supplementary Examination, Dec./Jan., 2022-2023

Degree	B. Tech.	**Program**	CSE, CSE (AI & ML) & CSE (DS)			**Academic Year**	2022 - 2023
Course Code	20CS305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	Computer Organization						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the need of a shift Register?	20CS305.1	L1
2	What is an Interrupt?	20CS305.2	L1
3	What is the main purpose of the RISC?	20CS305.3	L1
4	What is signed magnitude method?	20CS305.4	L1
5	What are Peripherals?	20CS305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Distinguish between fixed point representation and floating point representation.	6M	20CS305.1	L2
6 (b)	Define Decoder. Explain about 3 – to – 8 – line decoder.	6M	20CS305.1	L1
	OR			
7 (a)	Explain about the error detection codes.	6M	20CS305.1	L2
7 (b)	With a neat sketch explain about multiplexers.	6M	20CS305.1	L2
8	Explain arithmetic, logic and shift micro-operations.	12M	20CS305.2	L2
	OR			
9 (a)	Explain memory reference instructions with an example.	5M	20CS305.2	L2
9 (b)	Describe the various phases involved in the instruction cycle.	7M	20CS305.2	L2
10	What do you mean by addressing mode? Explain the addressing modes.	12M	20CS305.3	L2
	OR			
11 (a)	How address sequencing is done in micro programmed control?	6M	20CS305.3	L2
11 (b)	Explain about stack organization.	6M	20CS305.3	L2
12 (a)	Explain the hardware implementation of signed magnitude addition and subtraction.	8M	20CS305.4	L2
12 (b)	Draw and explain the flowchart for booth multiplication algorithm.	4M	20CS305.4	L2
	OR			
13 (a)	What is overflow and underflow in floating point arithmetic?	6M	20CS305.4	L2
13 (b)	Explain the difference between signed and unsigned division.	6M	20CS305.4	L2
14 (a)	Explain the methods employed for establishing priority using Daisy – Chaining priority.	6M	20CS305.5	L2
14 (b)	Describe the various components like input/output units, memory unit, control unit, arithmetic logic unit connected in the basic organization of a computer.	6M	20CS305.5	L2
	OR			
15 (a)	Explain the method of DMA transfer in a computer system.	5M	20CS305.5	L2
15 (b)	Explain the concept of virtual memory. Why it is significant?	7M	20CS305.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

Semester End Regular/Supplementary Examination, Dec./Jan., 2022-2023

Degree	B. Tech.	**Program**	CSE, CSE (AI & ML) & CSE (DS)			**Academic Year**	2022 - 2023
Course Code	20CS305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Computer Organization**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No. Questions (1 through 5) — Learning Outcome (s) — DoK

1 What is the need of a shift Register? — 20CS305.1 — L1

Definition-1M

Need- 1M

The shift registers are used for temporary data storage. The shift registers are also used for data transfer and data manipulation. The serial-in serial-out and parallel-in parallel-out shift registers are used to produce time delay to digital circuits.

2 What is an Interrupt? — 20CS305.2 — L1

Definition: 2M

Interrupts are the signals generated by the external devices to request the microprocessor to perform a task. so an interrupt in computer architecture is a signal that requests the processor to suspend its current execution and service the occurred

3 What is the main purpose of the RISC? — 20CS305.3 — L1

Definition-1 M

Purpose-1M

RISC (reduced instruction set computer) is a microprocessor that is designed to perform a smaller number of types of computer instructions so that it can operate at a higher speed. So the processor architecture that shifts the analytical process of a computational task from the execution or runtime to the preparation or compile time.

4 What is signed magnitude method? — 20CS305.4 — L1

Definition-2M

In the signed magnitude method number is divided into two parts: Sign bit and magnitude. Sign bit is 1 for negative number and 0 for positive number. Magnitude of number is represented with the binary form of the number.

5 What are Peripherals? — 20CS305.5 — L1

Definition- 2M

Peripheral devices are those devices that are linked either internally or externally to a computer. Peripherals read information from or write in the memory unit on receiving a command from the CPU. They are considered to be a part of the total computer system. As they require a conversion of signal values, these devices can be referred to as electromechanical and electromagnetic devices.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Distinguish between fixed point representation and floating point representation.	6M	20CS305.1	L2

Difference-6M

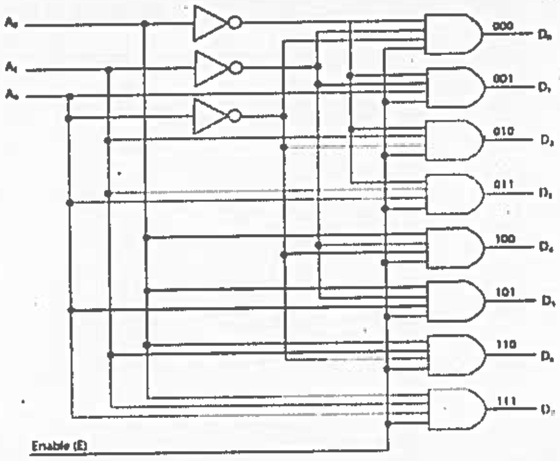
Fixed point representation	Floating point representation
Fixed point is a representation of real data type for a number that has a fixed number of digits after the radix point.	Floating point is a formulaic representation of real numbers as an approximation so as to support a tradeoff between range and precision.
While fixed point can be used to represent a limited range of values	floating point can be used to represent a wide range of values.
The performance of the fixed point is higher than floating point.	The performance of the fixed point is higher than floating point.
Floating point representation is more flexible than fixed point representation.	Floating point representation is more flexible than fixed point representation.
Higher Performance	Less Performance
Less flexibility	More flexible

6 (b)	Define Decoder. Explain about 3 – to – 8 – line decoder. Definition of Decoder- **2M** Explanation for 3 – to – 8 – line decoder & Diagram - **4 M**	6M	20CS305.1	L1

A decoder is a combinational circuit that modifies binary data from n input lines to a maximum of 2n unique output lines. An encoder creates the binary code corresponding to the input activated. A decoder gets a set of binary inputs and activates only the output that complements that input number.

<u>**3 – to – 8 – line decoder :**</u>The most preferred or commonly used decoders are n-to-m decoders, where m<= 2^n.An n-to-m decoder has n inputs and m outputs and is also referred to as an n * m decoder.

The following image shows a 3-to-8 line decoder with three input variables which are decoded into eight output, each output representing one of the combinations of the three binary input variables.



The three inverter gates provide the complement of the inputs corresponding to which the eight AND gates at the output generates one binary combination for each input.The most common application of this decoder is binary-to-octal conversion.

OR

7 (a) Explain about the error detection codes. 6M 20CS305.1 L2

Definition of Error detection codes- **2 M**
Explanation & Diagram -**4M**

Error-detecting codes are a sequence of numbers generated by specific procedures for detecting errors in data that has been transmitted over computer networks.

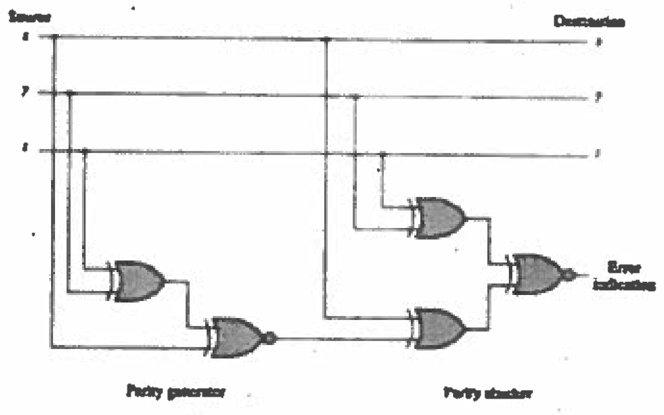
Parity Check

Parity check is done by adding an extra bit, called parity bit to the data to make number of 1s either even in case of even parity, or odd in case of odd parity.

While creating a frame, the sender counts the number of 1s in it and adds the parity bit in following way

- In case of even parity: If number of 1s is even then parity bit value is 0. If number of 1s is odd then parity bit value is 1.
- In case of odd parity: If number of 1s is odd then parity bit value is 0. If number of 1s is even then parity bit value is 1.

Figure 3-3 Error detection with odd parity bit.



7 (b) With a neat sketch explain about multiplexers. 6M 20CS305.1 L2

Definition of Multiplexers- **2 M**
Explanation & Diagram -**4M**

A Multiplexer (MUX) can be described as a combinational circuit that receives binary information from one of the 2^n input data lines and directs it to a single output line. The selection of a particular input data line for the output is decided on the basis of selection lines.

A 4-to-1-line multiplexer is shown in Fig. 2-4. Each of the four data inputs I0 through I, is applied to one input of an AND gate. The two selection inputs 51 and 50 are decoded to select a particular AND gate. The outputs of the AND gates are applied to a single OR gate to provide the single output. To demon strate the circuit operation, consider the case when 5150 = 10. The AND gate associated with input I, has two of its inputs equal to 1. The third input of the gate is connected to I2• The other three AND gates have at least one input equal to 0, which makes their outputs equal to 0. The OR gate output is now equal to the value of I,, thus providing a path from the selected input to the output.

The 4-to-1 line multiplexer of Fig. 2-4 has six inputs and one output. A truth table describing the circuit needs 64 rows since six input variables can have 2 6 binary combinations. This is an excessively long table and will not be shown here. A more convenient way to describe the operation of multiplexers is by means of a function table. The function table for the multiplexer is shown in Table 2-3. The table demonstrates the relationship between the four data inputs and the single output as a function of the selection inputs s1 and s0.

Figure 2-4 4-to-1-line multiplexer.

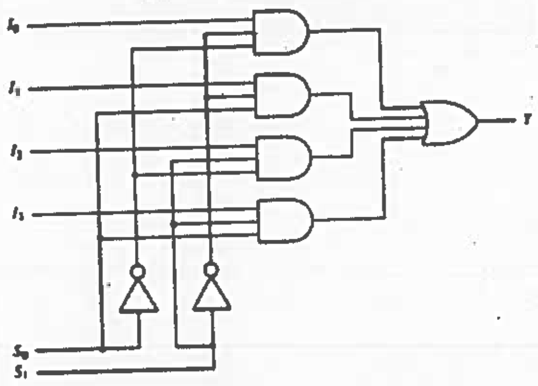


TABLE 2-3 Function Table for 4-to-1 Line Multiplexer

Select		Output
S_1	S_0	Y
0	0	I_0
0	1	I_1
1	0	I_2
1	1	I_3

8	Explain arithmetic, logic and shift micro-operations.	12M	20CS305.2	L2

Arithmetic Microoperations -4M
Logic Microoperations –Explanation-3M,Applications-1M
Shift Microoperations -4M

The basic arithmetic micro operations are addition, subtraction, increment, decrement, and shift. Arithmetic shifts are explained later in conjunction with the shift micro operations. The arithmetic micro operation defined by the statement R3 <-R1+R2.

Binary Adder: To implement the add micro operation with hardware, we need the registers that hold the data and the• digital component that performs the arithmetic addition. The digital circuit that forms the arithmetic sum of two bits and a previous carry is called a full-adder. The digital circuit that generates the arithmetic sum of two binary numbers of any lengths is called a binary adder. The below figure (c) shows the interconnections of four full-adders (FA) to provide a 4-bit binary adder..

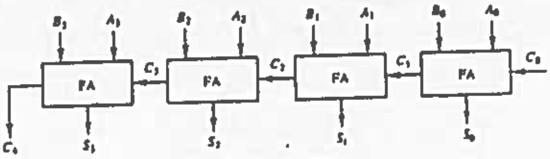


Figure (c): a 4 bit binary adder

Binary Adder-Subtractor The subtraction of binary numbers can be done most conveniently by means of complements. A - B can be• done by taking the 2's complement of B and added with A

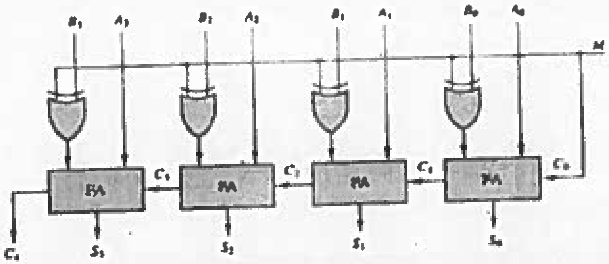


Figure (d): 4-bit adder-subtractor

Binary Incrementer: The increment micro operation adds one to a number in a register. For example, if a 4-bit registerhas a binary value 0110, it will go to 0111 after it is incremented.

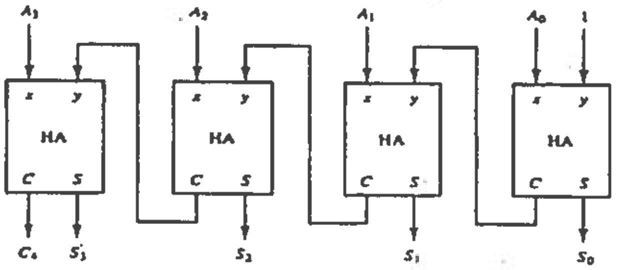


Figure (a): A 4-bit binary incrementer

LOGIC MICROOPERATIONS Logic micro operations specify binary operations for strings of bits stored in registers. These operations consider each bit of the register separately and treat them as binary variables. For example, the exclusive-OR micro operation with the contents of two registers R1 and R2is symbolized by the statement,

$$P:\ R1 \leftarrow R1 \oplus R2$$

List of Logic Microoperations: There are 16 different logic micro operations that can be performed with two binary variables

The 16 logic microoperations are derived from these functions by replacing variable x by the binary content of register A and variable y by the binary content of register B.

TABLE 4-6 Sixteen Logic Microoperations

Boolean function	Microoperation	Name
$F_0 = 0$	$F \leftarrow 0$	Clear
$F_1 = xy$	$F \leftarrow A \wedge B$	AND
$F_2 = xy'$	$F \leftarrow A \wedge \overline{B}$	
$F_3 = x$	$F \leftarrow A$	Transfer A
$F_4 = x'y$	$F \leftarrow \overline{A} \wedge B$	
$F_5 = y$	$F \leftarrow B$	Transfer B
$F_6 = x \oplus y$	$F \leftarrow A \oplus B$	Exclusive-OR
$F_7 = x + y$	$F \leftarrow A \vee B$	OR
$F_8 = (x + y)'$	$F \leftarrow \overline{A \vee B}$	NOR
$F_9 = (x \oplus y)'$	$F \leftarrow \overline{A \oplus B}$	Exclusive-NOR
$F_{10} = y'$	$F \leftarrow \overline{B}$	Complement B
$F_{11} = x + y'$	$F \leftarrow A \vee \overline{B}$	
$F_{12} = x'$	$F \leftarrow \overline{A}$	Complement A
$F_{13} = x' + y$	$F \leftarrow \overline{A} \vee B$	
$F_{14} = (xy)'$	$F \leftarrow \overline{A \wedge B}$	NAND
$F_{15} = 1$	$F \leftarrow$ all 1's	Set to all 1's

Logic microoperations are very useful for manipulating individual bits or a portion of a word stored in a register. They can be used to change bit values, delete a group of bits, or insert new bit values into a register.

1. selective-set (or) the OR microoperation
2. Selective-complement (or) the exclusive-OR microoperation.
3. selective-clear (or) the AB' microoperation.
4. Mask (or) AND microoperation

SHIFT MICROOPERATIONS

Shift micro operations are used for serial transfer of data. The contents of a register can be shifted to the left or the right. Determines the type of shift.

There are three types of shifts:

1. Logical shift
2. Circular shift
3. Arithmetic shift

OR

9 (a) Explain memory reference instructions with an example. 5M 20CS305.2 L2

Explanation- **5M**

MEMORY-REFERENCEINSTRUCTIONS:

AND : AND to AC This is an instruction that performs the AND logic operation on pairs of bits in AC and the memory word specified by the effective address. The result of the operation is transferred to AC.

AND : AND to AC This is an instruction that performs the AND logic operation on pairs of bits in AC and the memory word specified by the effective address.

LDA: Load to AC This instruction transfers the memory word specified by the effective address to AC.

STA: Store AC This instruction stores the content of AC into the memory word specified by the effective address.

BUN: Branch Unconditionally This instruction transfers the program to the instruction specified by the effective address.¬ The BUN instruction allows the programmer to specify an instruction out of sequenceandwe say¬ that the program branches (or jumps) unconditionally.

BSA: Branch and Save Return Address This instruction is useful for branching to a portion of the program called a subroutine or procedure. When executed, the BSA instruction stores the address of the next instruction in sequence (which is available in PC) into a memory location specified by the effective address. The effective address plus one is then transferred to PC to serve as the address of the first instruction in the subroutine.

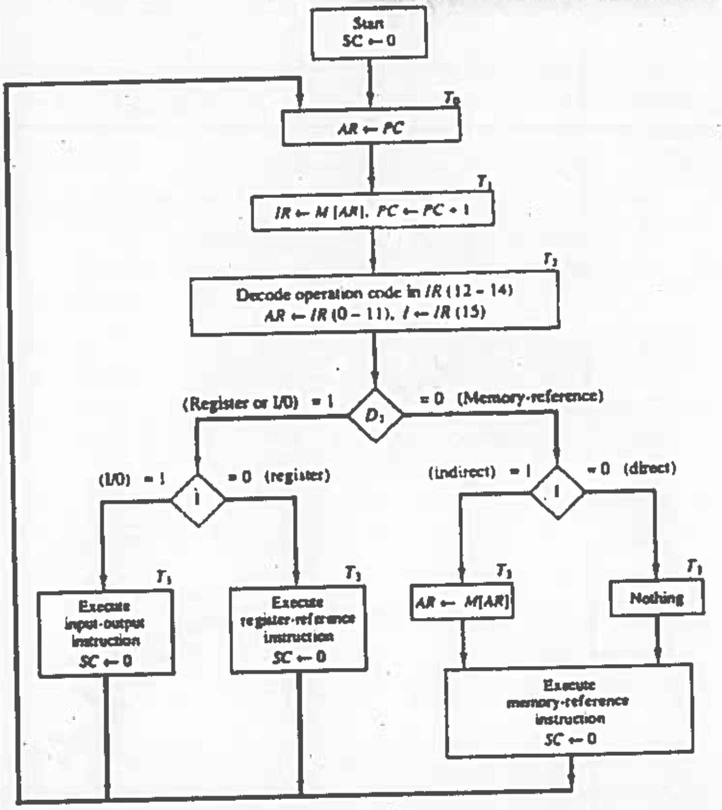
ISZ: Increment and Skip if Zero This instruction increments the word specified by the effective address, and if the incremented value is equal to 0, PC is incremented by 1.

9 (b)	Describe the various phases involved in the instruction cycle.	7M	20CS305.2	L2

Explanation-2 **M**
Fetch-1 **M**
Decode-1**M**
Read-1**M**
Execute-1**M**
Instruction Cycle flow chart-1**M**.

A program residing in the memory unit of the computer consists of a sequence of instructions. The program is executed in the computer by going through a cycle for each instruction. Each instruction cycle in turn is subdivided into a sequence of subcycles or phases. In the basic computer each instruction cycle consists of the following phases:

1. Fetch an instruction from memory.
2. Decode the instruction.
3. Read the effective address from memory if the instruction has an indirect address.
4. Execute the instruction.



10	What do you mean by addressing mode? Explain the addressing modes.	12M	20CS305.3	L2

Definition-1 M
Types of Addressing Modes: **11 M**

Addressing mode is the way in which the location of an operand can be specified in an instruction. It generates an effective address (the actual address of the operand). Instruction format with mode field
Types of Addressing Modes:

1. Implied Mode 2. Immediate Mode 3. Register Mode 4. Register Indirect Mode: 5. Autoincrement or Autodecrement Mode 6. Direct Address Mode 7. Indirect Address Mode 8. Relative Address Mode 9. Indexed Addressing Mode 10. Base Register Addressing Mode.

OR

11 (a)	How address sequencing is done in micro programmed control?	6M	20CS305.3	L2

Steps-**6M**

Microinstructions are stored in control memory in groups, with each group specifying a routine.

Step-1: An initial address is loaded into the control address register when power is turned on inthe computer.• This address is usually the address of the first microinstruction that activates the instruction fetch• routine. The fetch routine may be sequenced by incrementing the control address register through the rest of its• microinstructions. At the end of the fetch routine, the instruction is in the instruction register of the computer. 2 M

Step-2: The control memory next must go through the routine that determines the effective address of the• operand. A machine instruction may have bits that specify various addressing modes, such as indirect address and• index registers. The effective address computation routine in control memory can be reached through a branch microinstruction,• which is conditioned on the status of the mode bits of the instruction. When the effective address computation routine is completed, the address of the operand is available in the• memory address register. 2M

Step-3: The next step is to generate the microoperations that execute the instruction fetched frommemory.• The microoperation steps to be generated in processor registers depend on the operationcode part of the• instruction. Each instruction has its own micro-program routine stored in a given location of controlmemory.• The transformation from the instruction code bits to an address in control memory wherethe routine is located is• referred to as a mapping process. A mapping procedure is a rule that transforms the instruction code into a controlmemory address• 2 M

Step-4: Once the required routine is reached, the microinstructions that execute the instructionmay be sequenced by• incrementing the control address register. Micro-programs that employ subroutines will require an external register for storing thereturn address.• Return addresses cannot be stored in ROM because the unit has no writing capability.• When the execution of the instruction is completed, control must return to the fetchroutine.• This is accomplished by executing an unconditional branch microinstruction to the firstaddress of the fetch• routine. 2M

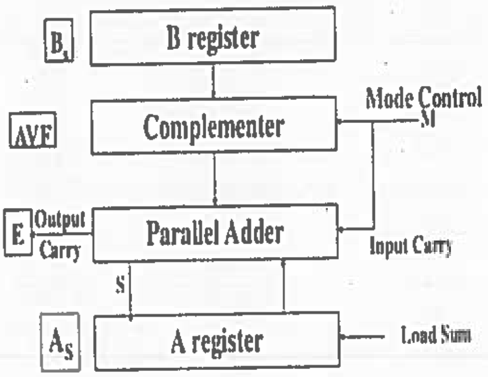
11 (b)	Explain about stack organization.	6M	20CS305.3	L2

Explanation: **3 M**
Daigram-**1M**
Push-**1M**
Pop-**1M**

STACK ORGANIZATION: A useful feature that is included in the CPU of most computers is a stack or last-in, first-out (LIFO) list. A stack is a storage device that stores information in such a manner that the item stored last is the first item retrieved. The operation of a stack can be compared to a stack of trays. The last tray placed on top of the stackϖ is the first to be taken off. The register that holds the address for the stack is called a stack pointer (SP) because its value always points at the top item in the stack. The two operations of a stack are the insertion and deletion of items. 1. Push or push-down (insertion operation) 2.

Pop or pop-up (deletion operation)

Hardware Algorithm for Signed-magnitude addition and subtraction



12 (a)	Explain the hardware implementation of signed magnitude addition and subtraction.	8M	20CS305.4	L2

Explanation – **6 M**

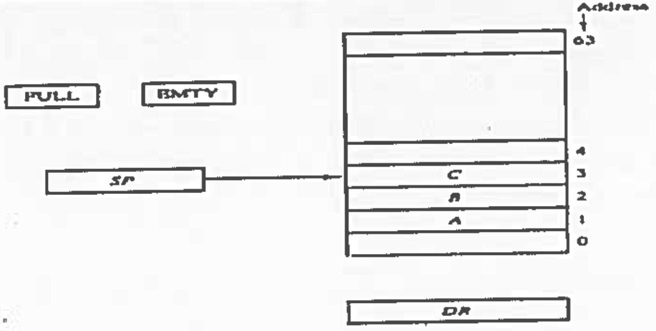
Daigram-**2 M**

Addition and Subtraction with Signed-Magnitude Data

The algorithms for addition and subtraction stated as follows (the words inside• parentheses should be used for the subtraction algorithm): Addition (subtraction) algorithm.• 1. when the signs of A and B are identical (different), add the two magnitudes and attach the sign of A to the result. 2. When the signs of A and B are different (identical), compare the magnitudes and 3. subtract the smaller number from the larger. Choose the sign of the result to be the same as A if A > B or the complement of the sign of A if A < B. If the two magnitudes are equal, subtract B from A and make the sign of the result positive.

Hardware Implementation for Addition and Subtraction with Signed-Magnitude Data

To implement the two arithmetic operations with hardware, it is first necessary that the two• numbers be stored in registers. Let A and B be two registers that hold the magnitudes of the numbers, and As, and Bs between flip-flops that hold the corresponding signs. Here parallel-adder is needed to perform the microoperation A + B. (Consists of Full adder). The complementer for generating the 2's Complement while performing subtraction operation.• (Consists of X-OR gate). Where M is Mode of operation. When M = 0, the output of B is transferred to the adder, the• input carry is 0, and the adder is equal to the sum A + B. When M =1, the l's complement of B is applied to the adder, the input carry is 1, and• output is = A + + 1. This is equal to A plus the 2's complement of B, which is equivalent to the A – B. Figure 1: Hardware for signed-magnitude addition and subtraction. A+ B. where EA is a register that•the magnitudes are added with a microoperation EA combines E and A. The value of E is transferred into the Add-overflow flip-flop AVF, if E is 1.• The magnitudes are subtracted by adding A to the 2's complement of B•



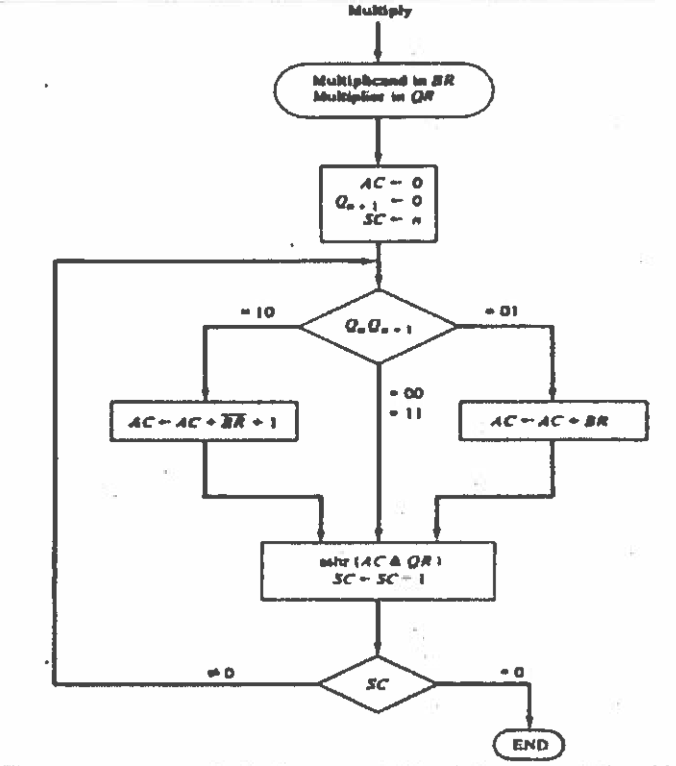
12 (b)	Draw and explain the flowchart for booth multiplication algorithm.	4M	20CS305.4	L2

Flow chart-**2 M**

Explanation-**2M**

Booth Multiplication Algorithm (for signed-2's complement numbers)

The flowchart for Booth algorithm is shown in below figure. AC and the appended bit Qn+1 are initially cleared to 0 and the sequence counter SC is set to a• number n equal to the number of bits in the multiplier. The two bits of the multiplier in Qn and Qn+1 are inspected.• If the two bits are equal to 10, it means that the first 1 in a string of 1's has been encountered.• This requires a subtraction of the multiplicand from the partial product in AC. If the two bits are equal to 01, it means that the first 0 in a string of 0's has been encountered.• This requires the addition of the multiplicand to the partial product in AC. When the two bits are equal, the partial product does not change. An overflow cannot occur• because the addition and subtraction of the multiplicand follow each other. The next step is to shift right the partial product and the multiplier (including bit Qn+1). This is an• arithmetic shift right (ashr) operation which shifts AC and QR to the right and leaves the sign bitin AC unchanged. The sequence counter is decremented and the computational loop is repeated n times.



OR

13 (a)	What is overflow and underflow in floating point arithmetic?	6M	20CS305.4	L2

Explanation of overflow-2 M

Example- 1M

Explanation of Underflow-2 M

Example- 1M

floating-point number in computer registers consists of two parts: a mantissa m and an exponent e. The two parts represent a number obtained from multiplying m times a radix r raised to the value of e ; thus m x re. The mantissa may be a fraction or an integer. The location of the radix point and the value of the radix r are assumed and are not included in the registers. For example, assume a fraction representation anda radix 10. The decimal number 537.25 is represented in a register with m = 53725 and e = 3 and is interpreted to represent the floating- point number 0.53725 x 10^3. A floating-point number is normalized if the most significant digit of the mantissa is nonzero. In this way the mantissa contains the maximumpossible number of significant digits. A zero cannot be normalized because it does not have a nonzero digit. It is represented in floating-point by all 0's in the mantissa and exponent.

A floating-point number that has a 0 in the most significant position of the mantissa is said to have an underflow. To normalize a number that contains an underflow, it is necessary to shift the mantissa to the left and decrement the exponent until a nonzero digit appears in the first position. In the example above, it is necessary to shift left twice to obtain .35000 x 10 3 . In most computers, a normalization procedure is performed after each operation to ensure that all results are in a normalized form.

When two normalized mantissas are added, the sum may contain an overflow digit. An overflow can be corrected easily by shifting the sum once to the right and incrementing the exponent. When two numbers are subtracted, the result may contain most significant zeros.

13 (b)	Explain the difference between signed and unsigned division.	6M	20CS305.4	L2

Signed Division-3 **M**
Unsigned Division-3 **M**

The Division of two fixed-point binary numbers in the signed-magnitude representation is done by the cycle of successive compare, shift, and subtract operations.Division of two fixed-point binary numbers in signed-magnitude representation is done with paper and pencil by a process of successive compare, shift, and subtract operations. Binary division is simpler than decimal division because the quotient digits are either 0 or 1 and there is no need to estimate how many times the dividend or partial remainder fits into the divisor. The division process is illustrated by a numerical example in Figure. The divisor B consists of five bits and the dividend A, of ten bits. The five most significant bits of the dividend are compared with the divisor. Since the 5- bit number is smaller than B, we try again by taking the six most significant bits of A and compare this number with B. The 6-bit number is greater than B, so we place a 1 for the quotient bit in the sixth position above the dividend. The divisor is then shifted once to the right and subtracted from the dividend. The difference is called a partial remainder because the division could have stopped here to obtain a quotient of 1 and a remainder equal to the partial remainder. The process is continued by comparing a partial remainder with the divisor. If the partial remainder is greater than or equal to the divisor, the quotient bit is equal to 1.

14 (a)	Explain the methods employed for establishing priority using Daisy – Chaining priority.	6M	20CS305.5	L2

Explanation-4 **M**

Daigram-2 **M**

The daisy-chaining method of creating priority includes a serial connection of all devices that request an interrupt. The device with the highest priority is located in the first position, followed by lower-priority devices up to the device with the lowest priority, which is situated last in the chain. This technique of connection between three devices and the CPU.

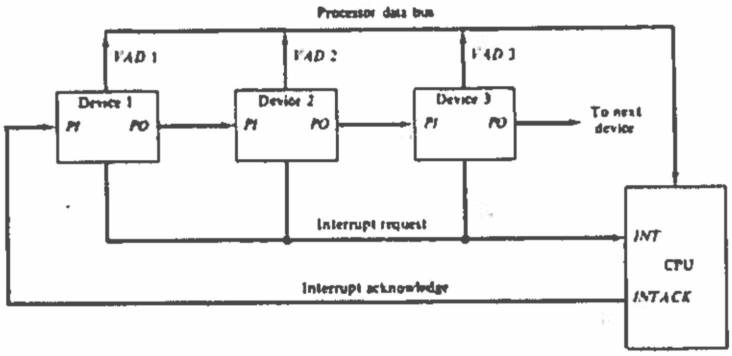


Figure 11-12 Daisy-chain priority interrupt.

14 (b)	Describe the various components like input/output units, memory unit, control unit, arithmetic logic unit connected in the basic organization of a computer.	6M	20CS305.5	L2

Input /Output Unit-**2 M**
Memory Unit- **1 M**
control unit – **1 M**
arithmetic logic unit-**1 M**
Diagram -**1 M**

Input /Output Unit : These components help users enter data and commands into a computer system. Data can be in the form of numbers, words, actions, commands, etc

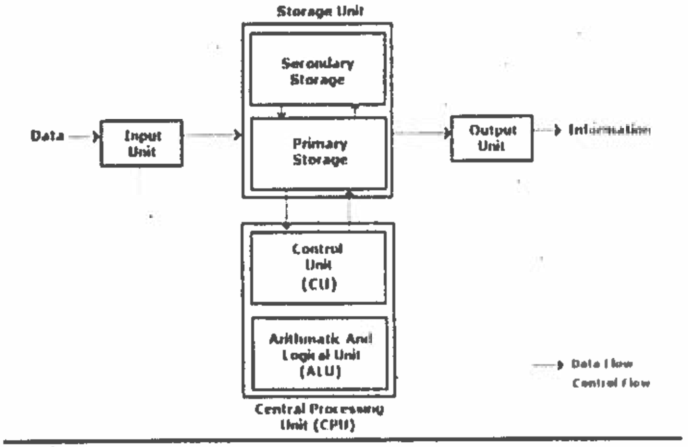
Memory Unit : Once a user enters data using input devices, the computer system stores this data in its memory unit.

control unit:
This unit is the backbone of computers. It is responsible for coordinating tasks between all components of a computer system.

Output Unit
The third and final component of a computer system is the output unit. After processing of data, it is converted into a format which humans can understand.

arithmetic logic unit
This part of the CPU performs arithmetic operations. It does basic mathematical calculations like addition, subtraction, division, multiplication, etc.

Block diagram of computer

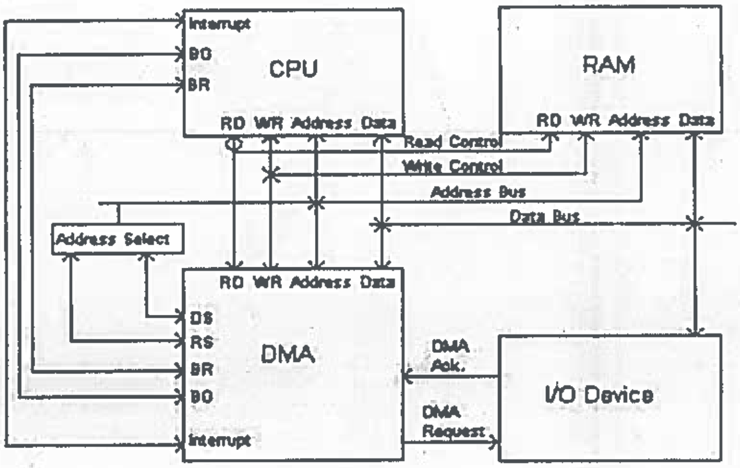


OR

15 (a)	Explain the method of DMA transfer in a computer system.	5M	20CS305.5	L2

DMA – 1M
Explanation- 3 M
Daigram-1 M

DMA basically stands for Direct Memory Access. It is a process which enables data transfer between the Memory and the IO (Input/ Output) device without the need of or you can say without the involvement of CPU during data transfer.



15 (b)	Explain the concept of virtual memory. Why it is significant?	7M	20CS305.5	L2

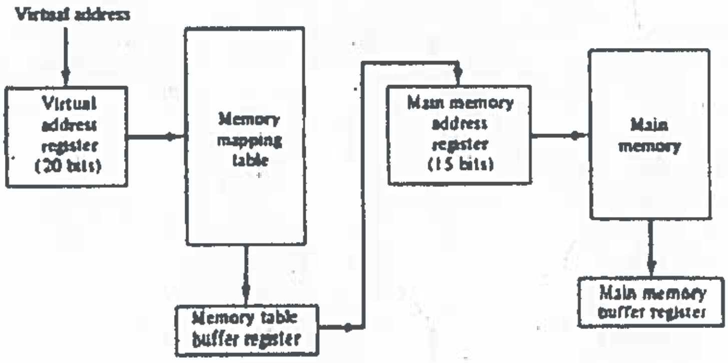
Explanation-4 M
Diagram- 2 M
Significance-1 M

Virtual memory is a memory management technique where secondary memory can be used as if it were a part of the main memory. Virtual memory is a common technique used in a computer's operating system (OS).

Significance :

- It can handle twice as many addresses as main memory.
- It enables more applications to be used at once.
- It frees applications from managing shared memory and saves users from having to add memory modules when RAM space runs out.

Figure 12-17 Memory table for mapping a virtual address.



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